National Curriculum (Vocational) Physical Science Level 4

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Contents

Subject Outcome I. Mechanics: Interpret, evaluate and apply motion of	
free-falling bodies	
Unit 1: Vertical projectile motion	3
<u>Dylan Busa</u>	
Unit 2: Graphs of vertical projectile motion	21
<u>Dylan Busa</u>	
Subject Outcome II. Mechanics: State, explain, interpret and apply	
momentum and force	
<u>Unit 1: Momentum</u>	45
<u>Dylan Busa</u>	
Unit 2: Collisions	67
<u>Dylan Busa</u>	
Subject Outcome III. Mechanics: Define, interpret and apply principles of	
work, power and energy	
Unit 1: Energy and work	89
Dylan Busa and Linda Pretorius	
<u>Unit 2: Energy-work theorem</u>	105
Dylan Busa and Linda Pretorius	
Unit 3: Mechanical power	119
<u>Dylan Busa and Linda Pretorius</u>	
Subject Outcome IV. Mechanics: State, evaluate and apply mechanical	
<u>advantage</u>	
Unit 1: Wheels and axles and their mechanical advantage	131
Leigh Kleynhans	
Unit 2: Pulleys and mechanical advantage	139
Leigh Kleynhans	
Unit 3: Gears and mechanical advantage	151
Leigh Kleynhans	

Subject Outcome V. Waves, sound and light: Describe and apply properties of waves in everyday life contexts

Unit 1: The Doppler effect	161
<u>Dylan Busa</u>	107
Unit 2: Electromagnetic radiation	183
<u>Dyian Busa</u>	107
Dylan Rusa	197
Dylan Busa	
Subject Outcome VI. Electricity and electronics: State, explain and apply	
principles used in electrodynamics	
Unit 1: Generating electricity	223
Leigh Kleynhans	
<u>Unit 2: Energy transfer in electrical circuits</u>	233
Leigh Kleynhans	
Subject Outcome VII. Electricity and electronics: State and apply principles	
and components used in electronics	
and components used in electronics	
Unit 1: Capacitance and inductance	245
Leigh Kleynhans	
Unit 2: Semi-conductors	251
Leigh Kleynhans	
Unit 3: Digital electronics	259
Leigh Kleynhans	
Subject Outcome VIII. Matter and Materials: Identify and critically evaluate	
<u>the impact of atomic nuclei on the quality of human, environmental and</u>	
socio-economic development	
Linit 1: Dadioactivity	275
	275
Subject Outcome IX. Matter and Materials: State, evaluate and apply	
properties of fluids on everyday life and industrial contexts	
Unit 1: Properties of Fluids	297
<u>Dylan Busa</u>	
<u>Subject Outcome X. Matter and Materials: Identify and apply knowledge of</u>	
organic molecules in everyday life and industrial context	
Unit 1: Organic Molecules	マンフ
	527

<u>Emma Harrage</u>

Unit 2: Naming organic molecules	347
Emma Harrage	
Unit 3: Physical Properties of Organic molecules	383
Emma Harrage	
Unit 4: Applications of organic molecules	397
Emma Harrage	
Unit 5: Plastics	405
Emma Harrage	
Subject Outcome XI. Chemical Change: Identify and apply knowledge of	
acids and bases	
Unit 1: Acids and Bases	423
Emma Harrage	
Subject Outcome XII. Chemical Change: Identify and apply knowledge of	
<u>electrochemical cells</u>	
Unit 1: Electrochemical Cells	453
Emma Harrage	
Subject Outcome XIII. Chemical Change: Describe, analyse and apply the	
rate and extent of chemical reaction	
Unit 1: Rates and extents of chemical reactions	491
Leigh Kleynhans	
Subject Outcome XIV. Chemical Systems and Industry: Identify and critically	
evaluate the impact of exploiting the lithosphere on the quality of human.	
environmental, and socio-economic development	
Unit 1: The Lithosphere	513
Emma Harrage	
Subject Outcome XV. Chemical Systems and Industry: Identify and critically	
evaluate the impact of scientific knowledge on the chemical industries and	
the quality of human, environmental and socio-economic	
Unit 1: Impact of Scientific Knowledge on Industry	553
Emma Harrage	

SUBJECT OUTCOME I MECHANICS: INTERPRET, EVALUATE AND APPLY MOTION OF FREE-FALLING BODIES

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Subject outcome

Subject outcome 2.1: Interpret, evaluate and apply motion of free-falling bodies



Learning outcomes

- Identify, analyse, describe and solve problems on vertical motion presented in diagrams, equations and graphs of motion.
- Identify that gravitational acceleration is independent of the mass of an object.
- Apply principles of projectile motion to diagrams and solve problems using equations of motion.



Unit 1 outcomes

By the end of this unit you will be able to:

- · Identify and describe vertical projectile motion.
- · Identify that gravitational acceleration is independent of the mass of an object.
- Identify, analyse and solve problems on vertical projectile motion using equations of motion.



Unit 2 outcomes

By the end of this unit you will be able to:

- Analyse and represent problems on projectile motion as graphs of motion.
- Identify, analyse and solve problems on projectile motion represented in and using graphs of motion.

Unit 1: Vertical projectile motion

DYLAN BUSA



By the end of this unit you will be able to:

- Identify and describe vertical projectile motion.
- Identify that gravitational acceleration is independent of the mass of an object.
- Identify, analyse and solve problems on vertical projectile motion using equations of motion.

What you should know

Before you start this unit, make sure you can:

- Describe motion and identify and define the components of motion including:
 - displacement
 - velocity
 - constant acceleration
 - Refer to level 2 subject outcome 2.1 units 1 and 2 if you need help with this.
- Analyse and calculate problems dealing with motion in one-dimension using equations of motion. Refer to <u>level 3 subject outcome 2.1 unit 1</u> if you need help with this.

Introduction

Consider the situation shown in figure 1. A big heavy bowling ball and a super light feather are held at a certain height above the ground. If they are dropped, what happens?



Figure 1: A bowling ball and a feather dropped from the same height

You have probably seen enough objects dropped to know that both the bowling ball and the feather will fall to the ground. But what makes them fall? Do they fall at the same speed? Which one hits the ground first?

Most, people when asked these questions, will respond that the objects fall because of gravity. Gravity 'pulls' the objects to the ground. They will go on to say that the bowling ball will fall faster and hit the ground first. This is after all our experience of the real world.

But, if the bowling ball falls faster and hits the ground first and if the force of gravity acting on the bowling ball and the feather is the same, then the bowling ball must fall faster because it is heavier. Its extra mass must cause it to be pulled towards the earth faster. Right? Well, no!

Note

If you have internet connection watch the video called Misconceptions About Falling Objects to see if these intuitions are correct.

Misconceptions About Falling Objects (Duration: 02.59)



Gravity

Many people believe that heavier objects fall faster because their experience shows them that bowling balls fall faster than feathers. But the only reason that bowling balls fall faster than feathers is because of air resistance. If you drop two similarly sized and shaped objects of very different masses from the same height, they will hit the ground **at the same time**!

If we make the objects the same shape and size, then we minimise the relative effects of air resistance on the two objects. Why not try this yourself now?



What to do:

Take a few pieces of paper and scrunch them up into a ball of about the same size as the real ball that you have. Make sure that each ball is the same size. Wrap a few pieces of tape around your paper ball to make sure it keeps its shape and size.

Now, hold both balls above your head and drop them. Do this a few times to make sure your results are repeatable.

What did you find?

You should have found, if both balls were the same shape and size, that they both hit the ground at the same time. It does not matter if one is heavier than the other. If we remove the relative effects of air resistance, we can see that both balls fall at the same rate and hit the ground at the same time.

To show that it does not matter what the mass of an object is, it will always fall at the same rate, watch the video called Brian Cox visits the world's biggest vacuum to see how, when all the air from a room is removed and the effects of air resistance are removed completely, a bowling ball and a feather fall at exactly the same rate and hit the ground at exactly the same time (see figure 2).

Brian Cox visits the world's biggest vacuum (Duration: 04:41)





Figure 2: A bowling ball and a feather fall at the same rate

We know from Newton's Law of Universal Gravitation that an object in the earth's gravitational field experiences a force pulling it towards the centre of the earth. If this is the only force acting on the object (i.e. we ignore or remove air resistance), then the object will accelerate towards the centre of the earth. We can show that the value of this acceleration is $9.8 \ {\rm m.s}^{-2}$. All objects, irrespective of their mass, experience the same rate of acceleration and, therefore, fall at the same rate.

For almost all everyday purposes, we can treat the acceleration due to gravity as a constant $9.8 \text{ m} \text{s}^{-2}$. It is only when one gets far away from the earth, that this 'gravitational force' changes.

Did you know?

In traditional *Newtonian physics* we speak about and treat gravity as a force. We say that the force of gravity on Earth accelerates objects at $9.8 m s^{-2}$.

However, *Einstein's theory of special relativity* explains that gravity is actually not a force at all. If you would like to find out why and what gravity really is watch the video called Why Gravity is NOT a Force.

Why Gravity is NOT a Force (Duration: 17.33)



Note

In all the problems that we will encounter in this unit and the next, we will ignore all other forces on the falling object, such as air resistance.

Motion due to gravity

In the next activity, we will investigate the motion of a ball as it is thrown vertically up into the air.



What to do:

Go outside where the air is as still as possible and throw a ball vertically up into the air. Watch carefully what happens to the ball between the moment it leaves your hand and when it hits the ground.

We will assign upwards as the positive direction.

- 1. Does the ball have an initial velocity?
- 2. Is this velocity increasing or decreasing?
- 3. Is the ball ever stationary and, if so, at what point?
- 4. If we ignore air resistance, what is the only force acting on the ball?
- 5. As the ball travels down, what is happening to its velocity?

What did you find?

1. As the ball leaves your hand travelling vertically upwards, it has a positive initial velocity.

- 2. This velocity is decreasing. As the ball gets higher and higher it slows down.
- 3. The ball's velocity slows down until, for a moment, it is stationary. This is the balls maximum height.
- 4. Ignoring air resistance, the only force acting on the ball from the moment it leaves your hand is the force of acceleration due to gravity which, because it acts down is -9.8 m.s^{-2} .
- 5. As the ball falls, its velocity increases. However, the ball's velocity is in the negative direction.

In activity 1.2 we dealt with an object (a ball in this case) that was thrown vertically up and fell vertically down and the only force the object experienced throughout was acceleration due to gravity. When objects travel vertically upwards and/or downwards while only experiencing acceleration due to gravity, we call this **vertical projectile motion**.

Activity 1.2 showed us a few important things about vertical projectile motion.



- Objects moving upwards or downwards in the earth's gravitational field, always accelerate **downwards** with this constant acceleration. Remember that non-zero acceleration means that the velocity is changing.
- Objects that have an initial velocity upwards will have **zero velocity at their greatest height** (h_{max}) . Remember that even though the velocity decreases to zero and then increases in the opposite direction, the acceleration experienced by the object is always constant.

Time symmetry

There is one aspect of free-falling bodies with initial upwards velocity that can be very helpful when solving problems and this is called **time symmetry**.

Consider an object thrown upwards from a vertical height, say h. We know that it will travel upwards with decreasing velocity (there is a constant downwards acceleration) until it stops and starts to fall back down to the ground (increasing in velocity in the opposite direction).

Note

We say that any object subject only to acceleration due to gravity is falling even if it initially travels upwards.

The time that it takes for the object to fall back down to the same height h is the same as the time taken for the object to reach its maximum height h_{max} from height h originally. This happens because the object experiences uniform acceleration throughout.

In addition, and again, because the object experiences a constant acceleration, the magnitude of the

velocity at which the object reaches height h on its way down is the same as the magnitude of the velocity with which the object was initially thrown upwards at height h. Obviously, the velocities are opposite in sign.



Equations of motion of falling objects

The equations of motion that you learnt about and used in levels 2 and 3 are equally applicable to the motion of falling objects. Why? These equations apply to objects moving in a straight line and with a constant acceleration. Falling objects (objects moving up or down vertically) move in a straight line with a constant acceleration. In all these cases we know that the acceleration is due to gravity. Therefore, we know that $a = g = 9.8 \text{ m.s}^{-2}$.

We often refer to the movement of objects vertically up or down as **vertical projectile motion**.

Here is a reminder of the equations of motion you should know.

- $\cdot v_f = v_i + at$
- $egin{array}{ccc} \cdot & s = v_i t + rac{1}{2} a t^2 \ \cdot & v_f{}^2 = {v_i}^2 + 2 a s \end{array}$

where, in the case of vertical projectile motion:

- v_i is the object's initial velocity $(m.s^{-1})$
- v_f is the object's initial velocity (m.s⁻¹)
- s is the object's vertical displacement (m) or its change in vertical position $(x_f x_i)$
- t is the time interval (s)
- a is the acceleration due to gravity ($a = g = 9.8 \text{ m.s}^{-2}$).

Take note!

Remember that when we use the equations of motion, we are dealing with vectors which have magnitude and direction. Therefore, you need to decide which direction will be the positive direction so that your vectors have the correct signs.



Example 1.1

A ball is thrown vertically upwards with an initial velocity of $7 m s^{-1}$ from a height of 1 m. Ignore air resistance.

- 1. What is the maximum height above the thrower's hand reached by the ball?
- 2. What is the maximum height above the ground reached by the ball?
- 3. How long does it take the ball to reach its maximum height?
- 4. How long after reaching its maximum height will the ball hit the ground?
- 5. How long after leaving the thrower's hand will the ball hit the ground?

Solutions

We are told that the ball is thrown vertically upwards and that we can ignore air resistance. Therefore, we are dealing with vertical projectile motion where the ball experiences a constant acceleration due to gravity. This is the only force acting on the ball.

1. We know that the ball's velocity when it reaches its maximum height will be 0 m.s^{-1} . Therefore, we know the ball's initial and final velocity and we can calculate its displacement. We will assign upwards as the positive direction.

$$v_i = 7 \text{ m.s}^{-1}$$

 $v_f = 0 \text{ m.s}^{-1}$
 $a = g = -9.8 \text{ m.s}^{-2}$
 $v_f^2 = v_i^2 + 2as$
 $\therefore 2as = v_f^2 - v_i^2$
 $\therefore s = \frac{v_f^2 - v_i^2}{2a}$
 $= \frac{(0 \text{ m.s}^{-1})^2 - (7 \text{ m.s}^{-1})^2}{2 \times (-9.8 \text{ m.s}^{-2})}$
 $= \frac{-49 \text{ m}^2 \cdot \text{s}^{-2}}{-19.6 \text{ m.s}^{-2}}$
 $= 2.5 \text{ m}$

Remember that upwards is positive and acceleration due to gravity always acts down

The ball's maximum height above the thrower's hand is $2.5\,$ m.

- 2. The ball was initially thrown from a height of 1 m above the ground. If its maximum height above the thrower's hand is 2.5 m, then its maximum height above the ground is 2.5 m + 1 m = 3.5 m.
- 3. We know that the ball's maximum height above the thrower's hand is 2.5 m and that its velocity at this height is 0 m.s^{-1} . Therefore, we have two equations we can use to calculate time. $v_t = v_t + at$

$$v_f = v_i + a \iota$$
 $s = v_i t + rac{1}{2} a t^2$

It will be simpler to use the first of these:

$$v_i = 7 \text{ m.s}^{-1}$$

$$v_f = 0 \text{ m.s}^{-1}$$

$$a = g = -9.8 \text{ m.s}^{-2}$$
Remember that upwards is positive and acceleration due to gravity always acts down
$$v_f = v_i + at$$

$$\therefore at = v_f - v_i$$

$$\therefore t = \frac{v_f - v_i}{a}$$

$$= \frac{0 \text{ m.s}^{-1} - 7 \text{ m.s}^{-1}}{-9.8 \text{ m.s}^{-2}}$$

$$= 0.71 \text{ s}$$

It will take 0.71 s for the ball to reach its maximum height.

4. We know from question 2 that the ball's maximum height above the ground is 3.5 m. We need to calculate how long it will take the ball to travel this distance.

$$\begin{split} v_i &= 0 \text{ m.s}^{-1} \\ s &= -3.5 \text{ m} \\ & \text{We need to be careful here.} \\ & \text{The change in the ball's } \\ & \text{vertical position is} \\ & x_f - x_i = 0 \text{ m} - 3.5 \text{ m} = -3.5 \text{ m} \\ a &= g = -9.8 \text{ m.s}^{-2} \\ & \text{Remember that upwards is} \\ & \text{positive and acceleration} \\ & \text{due to gravity always acts down} \\ s &= v_i t + \frac{1}{2}at^2 \\ & \text{But } v_i = 0 \text{ m.s}^{-1} \\ & \therefore s = \frac{1}{2}at^2 \\ & \therefore t^2 = \frac{2s}{a} \\ & \therefore t^2 = \frac{2s}{a} \\ & \therefore t = \sqrt{\frac{2s}{a}} \\ & = \sqrt{\frac{2 \times (-3.5 \text{ m})}{-9.8 \text{ m.s}^{-2}}} \\ &= 0.85 \text{ s} \\ & \text{It will take the ball } 0.85 \text{ s to reach the ground from its maximum height.} \end{split}$$

5. We know that the ball takes 0.71 s to reach its maximum height and then another 0.85 s to fall to the ground. Therefore, the total time is .0.71 s + 0.85 s = 1.56 s.

In example 1.1 we could have assigned downwards as the positive direction. All our answers would have been the same. Why not redo the example and assign downwards as the positive direction to check?



A cricketer hits a cricket ball so that it goes vertically upwards. If the ball takes to return to the initial height, determine its maximum height above the initial position.

Solution

In this question we will assign downwards as the positive direction.

We need to determine the maximum height of the ball. We are told that the ball takes 10 s after leaving the bat to rise to the maximum height and then to fall back down to the initial height. We know, given time symmetry, that the ball will spend the same amount of time travelling up to this height as it does to fall back down to this height. Therefore, the ball will take 5 s to get to its maximum height.

We also know, again given time symmetry, that its initial velocity on leaving the bat will be the same (but opposite) to its velocity as it falls back to the initial height.

Finally, we know that at its maximum height its velocity will be $0\,\,m.s^{-1}.$

If we focus on the portion of the ball's flight from its maximum height back to the initial height, we can use the fact that $v_i = 0$ m.s⁻¹.

$$v_i = 0 \text{ m.s}^{-1}$$

$$t = 5 \text{ s}$$

$$a = g = 9.8 \text{ m.s}^{-2}$$
Remember that downwards is positive and acceleration due to gravity always acts down
$$s = v_i t + \frac{1}{2}at^2$$
But $v_i = 0 \text{ m.s}^{-1}$

$$\therefore s = \frac{1}{2}at^2$$

$$= \frac{1}{2} \times 9.8 \text{ m.s}^{-2} \times (5 \text{ s})^2$$

$$= 122.5 \text{ m}$$

Because the ball travels 122.5 m to get back to its initial height from its maximum height, we know that the ball's maximum height above the bat is 122.5 m.

Note: The positive answer for displacement in this case tells us that the ball's change in position is downwards (it is falling) because we assigned downwards as positive.

It is important to note that in example 1.2 we did not calculate the ball's maximum height above the ground, just its maximum height above the bat. We do not know how high the bat is from the ground, and hence do not know the ball's maximum height above the ground.

This is important because it shows us that we can assign whatever coordinate system is convenient. In this case, we set our coordinate system to assign the zero position to the bat, not the ground.



- a. What height does the ball reach above the player's boot before it stops to fall back to the ground?
- b. How long does it take for the ball to reach its maximum height?
- 2. A child throws a toy vertically upwards so that the toy reaches a maximum height of $2\,\mathrm{m}$.
 - a. What initial velocity did the toy have?
 - b. How long does it take for the toy to reach its maximum height?
- 3. A tourist walking up the Leaning Tower of Pisa accidently drops her ice cream cone straight down. If the vertical distance covered in the last second is equal to the distance covered in first two seconds, find the height from which she dropped her ice-cream.



Figure 4: The Leaning Tower of Pisa

- 4. A scientist experimenting on a new mechanism of distributing food aid from airplanes drops a package from a hot-air balloon 60 m off the ground. At the moment of the drop, the balloon is ascending vertically upwards with a velocity of 5 m.s^{-1} .
 - a. With what velocity does the package hit the ground?
 - b. How long does the package take to hit the ground?

The <u>full solutions</u> can be found at the end of the unit.

Summary

In this unit you have learnt the following:

- Vertical projectile motion deals with objects that fall vertically due to gravity alone.
- Objects falling due to gravity experience a constant acceleration of $9.8\,\,{
 m m.s}^{-2}$ downwards.
- Because of this constant acceleration, objects fall with time symmetry, meaning that the time it takes an object to rise to its maximum height, is the same as the time it takes to fall back to its initial height.
- An object's velocity at its maximum height is zero.
- The velocity with which an object that is thrown vertically upwards will be the same but opposite to the velocity the object will have when it falls back to this initial position.
- The equations of motion can be used to solve vertical projectile problems.

Unit 1: Assessment

Suggested time to complete: 15 minutes

- 1. Two balls, A and B, are simultaneously thrown into the air from the same height above the ground. Ball A is thrown upwards, and ball B is thrown downwards with the same speed.
 - a. What can you say about the velocity of both balls as they hit the ground?
 - b. What can you say about the acceleration both balls experience?
- 2. What acceleration will a ball thrown vertically in the air experience at its maximum height?

Question 3 adapted from NC(V) Level 4 Physical Science Paper 1 November 2015 question 6

3. An object is projected vertically upwards at $_{6 m.s^{-1}}$ from the roof of a building. It strikes a concrete surface below after with a velocity of $_{14 m.s^{-1}}$. Ignore the effects of friction.



- a. How long will it take the object to reach its maximum height?
- b. Calculate the maximum height above the roof that the object reaches.
- c. Calculate the height of the building.

The <u>full solutions</u> can be found at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.

a. Let upwards be positive.

$$egin{aligned} v_i &= 15 \,\, {
m m.s^{-1}} \ v_f &= 0 \,\, {
m m.s^{-1}} \ a &= -9.8 \,\, {
m m.s^{-2}} \ v_f{}^2 &= v_i{}^2 + 2as \ dots &= rac{v_f{}^2 - v_i{}^2}{2a} \ dots &= 11.48 \,\, {
m m} \end{aligned}$$

b.

$$v_i = 15 ext{ m.s}^{-1}$$

 $v_f = 0 ext{ m.s}^{-1}$
 $a = -9.8 ext{ m.s}^{-2}$
 $v_f = v_i + at$
 $\therefore t = rac{v_f - v_i}{a}$
 $= 1.53 ext{ s}$

2.

a. Let upwards be positive

$$egin{aligned} s &= 2 \ {
m m} \ v_f &= 0 \ {
m m.s}^{-1} \ a &= -9.8 \ {
m m.s}^{-2} \ v_f{}^2 &= v_i{}^2 + 2as \ dots \ v_i{}^2 &= v_f{}^2 - 2as \ dots \ v_i &= \sqrt{v_f{}^2 - 2as} \ dots \ e &= 6.26 \ {
m m.s}^{-1} \end{aligned}$$

The ball had an initial velocity of $6.26\ m.s^{-1}$ upwards.

b.

$$v_i = 6.26 \text{ m.s}^{-1}$$

 $v_f = 0 \text{ m.s}^{-1}$
 $a = -9.8 \text{ m.s}^{-2}$
 $v_f = v_i + at$
 $\therefore t = rac{v_f - v_i}{a}$
 $= 0.64 \text{ s}$

3. Let upwards be positive.

First two seconds:
$$v_i = 0 \,\,{
m m.s}^{-1}$$

$$egin{aligned} v_i &= 0 ext{ m.s} \ t &= 2 ext{ s} \ s &= v_i t + rac{1}{2} a t^2 \ &= rac{1}{2} imes 4 imes a \ &= 2a \end{aligned}$$

Last second:

$$t=1 ext{ s}$$

 $s=v_it+rac{1}{2}at^2$
 $=v_i+rac{1}{2}a$
But $s=2a$
 $\therefore 2a=v_i+rac{1}{2}a$
 $\therefore v_i=rac{3}{2}a$

Last second:

 $egin{aligned} v_f &= v_i + at \ &= rac{3}{2}a + a \ &= rac{5}{2}a \end{aligned}$

We know that for the full fall: -1

$$v_i = 0 \text{ m.s}^{-1}$$

 $v_f = \frac{5}{2}a$
 $a = -9.8 \text{ m.s}^{-1}$
 $f_v^2 = v_i^2 + 2as$
 $\therefore s = \frac{f_v^2 - v_i^2}{2a}$
 $= \frac{\frac{25}{4}a^2}{2a}$
 $= \frac{25}{8}a$
 $= -30.625 \text{ s}$

Therefore, the ice-cream was dropped from a height of $30.625~\mathrm{m}$.

4.

a. Let upwards be positive.

To calculate the velocity with which the package hits the ground, we need to calculate the maximum height above the ground from which the package falls.

 $egin{aligned} v_i &= 5 \,\, {
m m.s}^{-1} \ v_f &= 0 \,\, {
m m.s}^{-1} \ a &= -9.8 {
m m.s}^{-2} \ v_f{}^2 &= v_i{}^2 + 2as \ dots \, s &= rac{v_f{}^2 - v_i{}^2}{2a} \ &= 1.28 \,\, {
m m} \end{aligned}$

Therefore, the maximum height of the package above the ground is $60\,\,\mathrm{m}+1.28\,\,\mathrm{m}=61.28\,\,\mathrm{m}.$

$$egin{aligned} v_i &= 0 \,\,\mathrm{m.s^{-1}} \ a &= -9.8\mathrm{m.s^{-2}} \ s &= -61.28 \,\,\mathrm{m} \ v_f{}^2 &= v_i{}^2 + 2as \ dots \, v_f &= \sqrt{v_i{}^2 + 2as} \ dots \, v_f &= \sqrt{v_i{}^2 + 2as} \ &= -34.66 \,\,\mathrm{m.s^{-1}} \end{aligned}$$

Therefore the package will have a velocity of $34.66\ {\rm m.s^{-1}}$ downwards when it hits the ground.

b.

$$egin{aligned} v_i &= 0 \,\, {
m m.s^{-1}} \ v_f &= -34.66 \,\, {
m m.s^{-1}} \ a &= -9.8 {
m m.s^{-2}} \ v_f &= v_i + at \ \ddots t &= rac{v_f - v_i}{a} \ &= 3.54 \,\, {
m s} \end{aligned}$$

Back to Exercise 1.1

Unit 1: Assessment

1.

- a. The velocity of both balls will be the same as they hit the ground. By the time ball A returns to the point from which it was thrown, its velocity will have the same magnitude with which it was initially thrown but in the downward direction. This will be the same velocity with which ball B was thrown.
- b. Both balls experience the same constant acceleration due to gravity $-9.8~{
 m m.s}^{-2}$ downwards.
- 2. $9.8\ m.s^{-2}$ downwards

3.

b.

a. Let upwards be positive.

$$egin{aligned} &v_i = 6 \,\,\mathrm{m.s^{-1}} \ v_f = 0 \,\,\mathrm{m.s^{-1}} \ a = -9.8 \mathrm{m.s^{-2}} \ v_f = v_i + at \ dots t = rac{v_f - v_i}{a} \ = 0.61 \,\,\mathrm{s} \ v_i = 6 \,\,\mathrm{m.s^{-1}} \ v_f = 0 \,\,\mathrm{m.s^{-1}} \ a = -9.8 \mathrm{m.s^{-2}} \ t = 0.61 \,\,\mathrm{s} \ s = v_i t + rac{1}{2} a t^2 \ = 1.84 \,\,\mathrm{m} \end{aligned}$$

The ball will reach a maximum height above the roof of $1.84~\mathrm{m}$.

c. The ball strikes the ground 5 s after being thrown. Therefore, it strikes the ground 5 s-0.61 s = 4.39 s after reaching its maximum height. Its velocity on hitting the ground is 14 m.s^{-1} .

$$egin{aligned} v_i &= 0 \,\, {
m m.s}^{-1} \ v_f &= -14 \,\, {
m m.s}^{-1} \ a &= -9.8 {
m m.s}^{-2} \ t &= 4.93 \,\, {
m s} \ v_f{}^2 &= v_i{}^2 + 2as \ dots \, s &= rac{v_f{}^2 - v_i{}^2}{2a} \ &= -10 \,\, {
m m} \end{aligned}$$

Therefore, the ball fell 10 m from its maximum height. But, its maximum height above the building was 1.84 m, therefore, the building is 10 m–1.84 m = 8.16 m high.

Back to Unit 1: Assessment

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Unit 2: Graphs of vertical projectile motion

DYLAN BUSA



Unit outcomes

By the end of this unit you will be able to:

- Analyse and represent problems on projectile motion as graphs of motion.
- Identify, analyse and solve problems on projectile motion represented in and using graphs of motion.

What you should know

Before you start this unit, make sure you can:

- Describe and identify vertical projectile motion. Refer to <u>unit 1</u> if you need help with this.
- Describe and apply gravitational potential energy. Refer to <u>level 2 subject outcome 2.3</u> if you need help with this.
- Use graphs to describe and interpret motion in one dimension including:
 - displacement-time graphs
 - velocity-time graphs
 - acceleration-time graphs.
 Refer to level 3 subject outcome 2.1 unit 2 if you need help with this.
- Calculate and interpret the gradient of:
 - displacement-time graphs to determine velocity
 - velocity-time graphs to determine acceleration.
- Calculate the area of velocity-time graphs to find displacement. Refer to <u>level 3 subject outcome 2.1 units 2</u> and <u>3</u> if you need help with this.

Introduction

In the previous unit, we learnt about vertical projectile motion and how we can analyse and solve problems about vertical projectile motion using the standard equations of motion. In this unit we will apply what we know about graphs of motion to analyse situations involving vertical projectile motion. Remember, that graphs of motion are just the graphical representations of the equations of motion.

Review graphs of motion

If you need to, now is a good time to revise what we learnt about graphs of motion in <u>level 3 subject outcome</u> 2.1 units 2 and 3. In level 3, we used these graphs of motion to analyse problems related to horizontal motion in one dimension.

Here is a summary of what you should know.



Figure 1: Summary of graphs of motion

Graphs of vertical projectile motion

Suppose you throw a ball directly up into the air. What do you think the displacement-time graph of this object's motion will look like as it rises into the air and then falls back down and lands in your hand? Draw a rough displacement-time graph for this motion before reading on. Assume that upwards is positive.

- 1. Initially (at t = 0 s), the ball will be in your hand and so, relative to your hand, it will have a displacement of 0 m.
- 2. After leaving your hand, the ball travels up and away from your hand while experiencing a constant negative acceleration due to gravity.
- 3. Eventually, the ball will reach its maximum height (its maximum displacement from your hand).
- 4. The ball will then start to fall while continuing to experience the same constant negative acceleration due to gravity.
- 5. Finally, the ball will land back in your hand and so have a final displacement, relative to your hand, of 0 m.

If we were to draw this motion, it would look like the displacement-time graph in figure 2.



Figure 2: Displacement-time graph of a ball thrown vertically in the air and caught again

Notice how the slope is not constant. The ball is experiencing a constant negative acceleration due to gravity. On the way up, the ball is decelerating (the slope is decreasing). At its maximum height, the velocity of the ball is zero (the slope is zero). After this, the ball starts to accelerate in the negative direction and so the slope of the line is negative and increasing.

Note that in the graph in figure 2, the displacement is relative to your hand. If your hand were some distance above the ground and we plotted the same motion relative to the ball's displacement from the ground, the displacement-time graph would look like figure 3.



Figure 3: Displacement-time graph of a ball thrown vertically in the air and caught again with reference to the ground

Now, consider the same situation (a ball thrown vertically up and caught again) but this time draw the velocity-time graph of this motion. Again, make your own sketch before reading on. Again, assume that upwards is positive.

- 1. Initially (at t = 0 s), the ball will be travelling with some initial positive velocity.
- 2. The ball decelerates while travelling up because it is experiencing a constant negative acceleration due to gravity.
- 3. Eventually, the ball will reach its maximum height at which its velocity will be 0 m.s^{-1} .
- 4. The ball will then start to accelerate in the negative direction as it falls because it continues to experience the same constant negative acceleration due to gravity.
- 5. Finally, the ball will land back in your hand with the same magnitude of velocity with which it was thrown (time symmetry).

If we were to draw this motion, it would look like the velocity-time graph in figure 4.



Figure 4: Velocity-time graph of a ball thrown vertically in the air and caught again

Remember that the final velocity is the same but opposite to the initial velocity and that the time it takes for the ball to travel from your hand to its maximum height is the same as the time it takes for the ball to fall from its maximum height back to your hand.

The slope of the line is constant because the ball experiences a constant acceleration due to gravity. The slop of the line is negative because we defined upwards as positive and acceleration due to gravity is always downwards. If we measured the slope of the velocity-time graph, it would be -9.8 m.s^{-2} .

Finally, draw the acceleration-time graph of this situation. This is the simplest graph to draw if you remember that the ball experiences a constant acceleration due to gravity. Again, assume that upwards it positive.

If we were to draw this motion, it would look like the acceleration-time graph in figure 5.



Figure 5: Acceleration-time graph of a ball thrown vertically in the air and caught again

The defining feature of vertical projectile motion is that the object is always subject to a constant acceleration due to gravity of $g = 9.8 \ m. s^{-2}$. Because we defined upwards as positive and because acceleration due to gravity always acts downwards, we represent this constant gravity as $-9.8 \ m. s^{-2}$ in our acceleration-time graph.

Drawing graphs of vertical projectile motion

Whenever you are asked to draw a graph representing vertical projectile motion, remember that:

- $\cdot\,\,$ the object experiences a constant acceleration due to gravity of $9.8\,\,{
 m m.s}^{-2}$; and
- this constant acceleration is always downwards.



Example 2.1

A learner stands on a first-floor balcony and throws a cricket ball vertically up into the air with an initial velocity of 4 m.s^{-1} . It reaches its maximum height and then falls to the ground. If the ball is thrown from a height of 10 m, draw an accurate:

- 1. displacement-time graph
- 2. velocity-time graph
- 3. acceleration-time graph.

Solutions

1. We know that the ball's initial displacement from the ground is 10 m. We will draw the

displacement-time graph relative to the ground and assign upwards as positive.

We need to calculate the following to draw the displacement-time graph:

- \cdot the time it takes for the ball to reach its maximum height
- the ball's maximum height above the ground
- the time it takes the ball to fall to the ground.

The time it takes for the ball to reach its maximum height:

$$egin{aligned} v_i &= 4 \ {
m m.s^{-1}} \ v_f &= -14 \ {
m m.s^{-1}} \ a &= -9.8 {
m m.s^{-2}} \ v_f &= v_i + at \ dots &= v_i + at \ dots &= \dfrac{v_f - v_i}{a} \ &= 0.41 \ {
m s} \end{aligned}$$

The ball's maximum height above the ground.

$$egin{aligned} v_i &= 4 \,\,{
m m.s^{-1}}\ a &= -9.8 \,\,{
m m.s^{-2}}\ t &= 0.41 \,\,{
m s}\ s &= v_i t + rac{1}{2} a t^2\ &= 0.82 \,\,{
m m} \end{aligned}$$

This is the height of the ball above the hand that threw it. Therefore, the ball's height above the ground is 0.82 m + 10 m = 10.82 m.

 $\mathbf{s} \boldsymbol{\cdot}$

The time it takes the ball to fall to the ground:

$$v_i = 0 \text{ m.s}^{-1}$$

$$a = -9.8 \text{-m.s}^{-2}$$

$$s = -10.82 \text{ m}$$
Remember that displacement is
$$x_f - x_i = 0 \text{ m} - 10.82 \text{ m} = -10.82 \text{ m}$$

$$s = v_i t + \frac{1}{2} a t^2$$
But $v_i = 0 \text{ m.s}^{-1}$

$$\therefore s = \frac{1}{2} a t^2$$

$$\therefore t = \sqrt{\frac{2s}{a}}$$

$$= 1.49 \text{ s}$$
Therefore, the total time the ball is in the air is $0.41 \text{ s} + 1.49 \text{ s} = 1.9$

Now we can draw an accurate displacement-time graph as follows.





Question adapted from Everything Science Grade 12 Worked example 6 page 93

A cricketer hits a cricket ball from the ground and the following graph of velocity vs time was drawn. Upwards was taken as positive. Study the graph and follow the instructions below:


- 2. Draw a sketch graph of the corresponding displacement-time graph. Label the axes.
- 3. Draw a sketch graph of the corresponding acceleration-time graph. Label the axes.

Solutions

- 1. If we look at the velocity-time graph we can see that the slope is negative. This immediately tells us that down is negative and up is positive. The ball starts off with a positive velocity of 19.6 m.s^{-1} . Therefore, the ball initially travels up in the air. After 2 s, the velocity of the ball is 0 m.s^{-1} . Therefore, at 2 s the ball reaches its maximum height and starts to fall down, travelling with an increasing but negative velocity. At 4 s, the ball's velocity is -19.6 m.s^{-1} (the same magnitude as its initial velocity) and so the ball is at the same height as where it started from, in this case the ground. The ball has returned to its starting position.
- 2. The only additional piece of information we need to draw the displacement-time graph is the ball's maximum height above the ground which we know is reached after $2_{\rm S}$. We can calculate this using an equation of motion.

$$egin{aligned} v_i &= 19.6 \,\,\mathrm{m.s^{-1}} \ a &= -9.8 ext{-m.s^{-2}} \ t &= 2 \,\,\mathrm{s} \ s &= v_i t + rac{1}{2} a t^2 \ &= 19.6 \,\,\mathrm{m} \end{aligned}$$

Alternatively, we can calculate the area under the given velocity-time graph between $t=0~{
m s}$ and $t=2~{
m s}$.

$$egin{aligned} Area &= rac{1}{2} imes b imes h \ &= rac{1}{2} imes 2 ext{ s} imes 19.6 ext{ m.s}^{-1} \ &= 19.6 ext{ m} \end{aligned}$$

We can see from the velocity-time graph that the total area under the graph between t = 0 s and t = 4 s is 0 m. In other words, this is confirmation that the ball's total displacement is 0 m and that the ball returns back to its starting point.





Exercise 2.1

- 1. The velocity-time graph below is of a bullet fired directly upwards from a gun. Downwards is assigned as the positive direction.
 - a. Describe the motion of the bullet.
 - b. Draw the corresponding displacement-time graph.



c. Acceleration-time

The <u>full solutions</u> can be found at the end of the unit.

Summary

In this unit you have learnt the following:

• Graphs for vertical projectile motion are similar to graphs for horizontal motion at constant acceleration.

Unit 2: Assessment

Suggested time to complete: 30 minutes

Question 1 adapted from NC(V) Level 4 Physical Science Paper 1 November 2019 question 6

1. A hot-air balloon is rising vertically at constant velocity. When the balloon is at a height of 88 m above the ground, a stone (mass 200 g) is dropped from the balloon. The displacement-time graph below represents the motion of the stone from the moment it is released from the balloon until the stone strikes the ground. Ignore the effect of air resistance.



Use the information supplied in the graph to answer the following questions.

- a. Calculate the gravitational potential energy of the stone the instant it is dropped from the balloon.
- b. State the velocity of the stone when it reaches its maximum height above the ground.
- c. Calculate the velocity of the stone the instant it strikes the ground.
- d. Calculate the velocity of the hot-air balloon the instant the stone is dropped from the balloon.
- e. Draw a rough acceleration versus time graph for the motion of the stone.
- 2. The velocity-time graph below shows the motion of an object. Upwards is the positive direction.



- a. Describe the motion of this object.
- b. Plot the corresponding displacement-time graph.
- c. Plot the corresponding acceleration-time graph.

The <u>full solutions</u> can be found at the end of the unit.

Unit 2: Solutions

Exercise 2.1

٦.

- a. The initial velocity of the bullet is -180 m.s^{-1} . This is upwards (in the negative direction). The bullet's velocity decreases due to the constant downwards (positive) acceleration due to gravity, until after a certain time its velocity is zero. This is the point where it reaches its maximum height. Thereafter, also because of the constant positive acceleration due to gravity, the bullet starts to fall back to the ground. As it does so, its velocity increases, until at 36.74 s, its velocity is 180 m.s^{-1} , the same magnitude with which it was initially fired. This means that the bullet is at the same height from which it was fired and its total displacement is 0 m.
- b. The velocity-time graph shows time symmetry. Therefore, we know that the time after which the bullet decelerates to $_{0\ m.s^{-1}}$ is exactly half the time it takes to achieve a velocity of $_{180\ m.s^{-1}}$.

Therefore, the bullet reaches its maximum height after $\frac{36.74 \text{ s}}{2} = 18.37 \text{ s}$.

The maximum height of the bullet:

$$egin{aligned} &v_i = 180 \,\,{
m m.s}^{-1}\ &a = -9.8\,{
m rm.s}^{-2}\ t = 18.37\,\,{
m s}\ &s = v_i t + rac{1}{2}at^2\ &= -1\,\,653\,\,{
m m} \end{aligned}$$

Note: The displacement is negative because we assigned downwards as positive as the bullet has travelled upwards.



c. Draw the corresponding acceleration-time graph.



2.

a. Let upwards be positive.

Velocity with which the ball first strikes the ground:

$$egin{aligned} v_i &= 0 \,\,{
m m.s^{-1}} \ a &= -9.8\,{
m rm.s^{-2}} \ s &= -5 \,\,{
m m} \ v_f{}^2 &= v_i{}^2 + 2as \ dots \, v_f &= \sqrt{v_i{}^2 + 2as} \ = -9.9 \,\,{
m m.s^{-1}} \end{aligned}$$

The time taken for the ball to first strike the ground

$$v_i = 0 \text{ m.s}^{-1}$$

 $v_f = -9.9 \text{ m.s}^{-1}$
 $a = -9.8 \text{-m.s}^{-2}$
 $v_f = v_i + at$
 $\therefore t = rac{v_f - v_i}{a}$
 $= 1.01 \text{ s}$

The initial velocity after the first bounce is ${9.9~{
m m.s}^{-1}\over 2}=4.95~{
m m.s}^{-1}\cdot$

The max height reached by the ball after its first bounce:

$$egin{aligned} v_i &= 4.95 \ {
m m.s}^{-1} \ v_f &= 0 \ {
m m.s}^{-1} \ a &= -9.8 { imes} {
m m.s}^{-2} \ v_f{}^2 &= v_i{}^2 + as \ dots &= rac{v_f{}^2 - v_i{}^2}{2a} \ &= 1.25 \ {
m m} \end{aligned}$$

The time taken for the ball to reach maximum height after its first bounce:

$$egin{aligned} v_i &= 4.95 \,\, {
m m.s}^{-1} \ v_f &= 0 \,\, {
m m.s}^{-1} \ a &= -9.8 {\sim} {
m m.s}^{-2} \ v_f &= v_i + at \ dots &= v_i + at \ dots &= rac{v_f - v_i}{a} \ &= 0.5 \,\, {
m s} \end{aligned}$$

Therefore, the total elapsed time is 1.01 $\mathrm{s}+0.5$ $\mathrm{s}=1.51$ s.

The velocity with which the ball first strikes the ground after the first bounce, due to time symmetry, is -4.95 m.s^{-1} . Therefore, the initial velocity after the second bounce is $\frac{4.95 \text{ m.s}^{-1}}{2.48 \text{ m.s}^{-1}}$. The ball bounces a second time after 1.51 s + 0.5 s = 2.01 s.

The max height reached by ball after the second bounce:

$$v_i = 2.48 \text{ m.s}^{-1}$$

 $v_f = 0 \text{ m.s}^{-1}$
 $a = -9.8 \text{-m.s}^{-2}$
 $v_f^2 = v_i^2 + 2as$
 $\therefore s = rac{v_f^2 - v_i^2}{2a}$
 $= 0.31 \text{ m}$

The time taken for the ball to reach its maximum height after the second bounce:

$$v_i = 2.48 \text{ m.s}^{-1}$$

 $v_f = 0 \text{ m.s}^{-1}$
 $a = -9.8 \text{-m.s}^{-2}$
 $v_f = v_i + at$
 $\therefore t = rac{v_f - v_i}{a}$
 $= 0.25 \text{ m}$

Therefore, the total elapsed time is 2.01 s + 0.25 s = 2.26 s.



c. Acceleration vs time



Note: When the ball bounces, it experiences a moment of great positive acceleration. We do not have enough information to determine what this is but for these very brief moments, the acceleration experienced by the ball is not that due to gravity. Therefore, we need to indicate this using open dots on the acceleration-time graph at these points in time.

Back to Exercise 2.1

Unit 2: Assessment

1.

а.

 ${
m Gravitational\ potential\ energy}=mgh$

 $= 0.2 \text{ kg} imes 9.8 \text{ m.s}^{-2} imes 88 \text{ m}$ = 172.48 J

- b. $0 \ \mathrm{m.s^{-1}}$
- c. The displacement-time graph indicates that upwards is positive.

 $egin{aligned} &v_i = 0 \,\, {
m m.s}^{-1} \ &t = 6.08 \,\, {
m s} - 1.56 {
m s} = 4.52 {
m s} \ &a = -9.8 {
m ~ m.s}^{-2} \ &v_f = v_i + at \ &= -44.30 \,\, {
m m.s}^{-1} \end{aligned}$

The stone's velocity on hitting the ground is $44.30\ m.s^{-1}$ downwards.

d. The stone travels upwards for $1.56~{
m s}$ before reaching its maximum height and having a velocity of

 $\begin{array}{l} 0 \ {\rm m.s}^{-1} \cdot \\ v_i = 0 \ {\rm m.s}^{-1} \\ t = 1.56 \ {\rm s} \\ a = -9.8 {\rm \cdot m.s}^{-2} \\ v_f = v_i + at \\ \therefore v_i = v_f - at \\ = 15.29 \ {\rm m.s}^{-1} \end{array}$

The velocity of the balloon at the moment the stone is released is $15.29~{
m m.s}^{-1}$ upwards.



Note

The motion of the stone in question 1 above may seem strange. How can a stone that is dropped travel upwards for a time before falling to the ground? This is possible only because the stone, at the moment of release, is travelling upwards with the hot-air balloon. It has an initial upwards velocity. As soon as the stone is released, acceleration due to gravity is the only force acting on the stone and it immediately begins to slow down. From the perspective of someone in the balloon, the stone will, therefore, seem to move downwards immediately. This is not the case, as the stone will keep moving up for a time but with an ever-decreasing velocity until it reaches its maximum height above the ground where its velocity is zero and it starts to fall towards the ground.

A very similar phenomenon happened when a Space Shuttle separated from its booster rockets (see figure 6). At the moment of separation, the booster rockets were travelling at the same speed as the Shuttle (about 1 $_{280}$ m.s⁻¹) and were 45 km above the ground. However, because of their initial upwards velocity, the booster rockets continued to rise and reached a maximum height of 67 km before starting to fall back to earth.



2.

a. The object has an initial velocity of 30 m.s^{-1} upwards but is decelerating. After 3.06 s, its velocity is 0 m.s^{-1} and then becomes negative indicating that at 3.06s s, the object reaches its maximum height and starts to accelerate downwards under gravity.

After another 3.06 s (a total time of 6.12 s) the object has achieved a negative (downwards) velocity of -30 m.s^{-1} indicating that it is back at its original height. At this point, its velocity instantaneously changes to being 30 m.s⁻¹ again indicating that the object has bounced and travelled upwards again. After another 3.06 s (a total time of 9.18 s), the object has once again reached its maximum height and its velocity is again 0 m.s⁻¹. The object starts to fall again and after another 3.06 s (total time of 12.24 s) it again has a downwards velocity of -30 m.s^{-1} having reached its starting point again.

b. Maximum height reached by the object:

$$v_i = 30 ext{ m.s}^{-1}$$

 $v_f = 0 ext{ m.s}^{-1}$
 $a = -9.8 ext{-m.s}^{-2}$
 $v_f^2 = v_i^2 + 2as$
 $\therefore s = rac{v_f^2 - v_i^2}{2a}$
 $= 45.92 ext{ m}$



Back to Unit 2: Assessment

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SUBJECT OUTCOME II MECHANICS: STATE, EXPLAIN, INTERPRET AND APPLY MOMENTUM AND FORCE

Subject outcome

Subject outcome 2.2: State, explain, interpret and apply momentum and force



 (\forall)

Learning outcomes

- Define and calculate momentum and calculate the change of momentum of an object.
- Define impulse $(F_{nett} \cdot t)$ as the change in momentum and use $F_{nett} \cdot \Delta t = m \cdot \Delta v$ for calculations and problem solving.
- Define Newton's third law and identify examples of interacting objects that exert equal forces on each other.
- Define the law of conservation of momentum.
- Identify and distinguish between elastic and inelastic collisions; and apply law using calculations.



Unit 1 outcomes

By the end of this unit you will be able to:

- · Define and calculate momentum.
- Calculate the change of momentum of an object.
- Define impulse $(F_{nett} \cdot t)$ as the change in momentum and use $F_{nett} \cdot \Delta t = m \cdot \Delta v$ for calculations and problem solving.



Unit 2 outcomes

By the end of this unit you will be able to:

- Define Newton's third law and identify examples of interacting objects that exert equal forces on each other.
- Define the law of conservation of momentum.

- Distinguish between elastic and inelastic collisions.
- \cdot Apply the law of conservation of momentum in calculations on different types of collisions.

Unit 1: Momentum

DYLAN BUSA



What you should know

Before you start this unit, make sure you can:

- Define and work with vector quantities. Refer to <u>level 2 subject outcome 2.1 unit 3</u> if you need help with this.
- Identify and apply principles of force. Refer to <u>level 2 subject outcome 2.2</u> if you need help with this.
- Describe, analyse and apply principles of concurrent forces. Refer to <u>level 3 subject outcome 2.2</u> if you need help with this.

Introduction

What would you rather be hit by a tennis ball travelling at 145 km/h or a cricket ball travelling at 50 km/h? In considering this, you may think that the tennis ball would be the better option because it is lighter and softer. However, it is travelling at about three times the speed.

Our experience of the real world and intuition tell us that heavier things hit harder. When a fly lands on your arm you hardly notice it. When a dog jumps on you, you feel it. The fly is light, and the dog is heavy.

However, our experience and intuition also tell us that faster things hit harder. A car travelling at 10 km/h is going to do far less damage if driven into a wall than the same car travelling at 100 km/h driven into the same wall.

Which property, the mass of an object or its velocity, matters more when considering the force it can exert?

Momentum

Moving objects can exert a force on other objects as a result of their momentum. Momentum can be simply defined as mass in motion. Every moving object, from stars to electrons, has momentum. If it is not moving, it has no momentum.

We define momentum (p) as the mass of an object multiplied by its velocity. Therefore, $p = m \cdot v$.

As velocity is a vector, momentum is also a vector. Whenever we calculate momentum, we need to note its direction.

This means that momentum is directly proportional to both mass and velocity. If you increase either mass or velocity, you increase momentum. If you double an object's velocity, you double its momentum. If you double an object's mass, you double its momentum. If you double both an object's velocity and mass, you quadruple its momentum.

As mass in measured in kg and velocity is measured in $m.s^{-1}$, momentum has units of kg.m.s⁻¹.



momentum? Assume that the moon's orbit is circular.

Solution

Step 1: Identify the information that has been given

- + The mass of the moon: $7.35 imes 10^{22}~{
 m kg}$
- The distance of the moon from earth: 384 400 $\,\mathrm{km}=384$ 400 $\,\mathrm{000}\,\mathrm{m}$
- + The duration of orbit: 27.3 days = 27.3 \times 24 hrs/day \times 60 min/hr \times 60 s/min = 2 358 720 s

We need to calculate the magnitude of the momentum of the moon. We do not need to worry about its direction in this case. We are told to assume that the moon's orbit is circular.

Step 2: Do the calculation

If the moon's orbit is circular, then we can assume that its distance from the earth is constant and equal to the radius of this orbit. With the orbital radius, we can calculate the orbital circumference. Circumference $= 2\pi r = 2 \times \pi \times 384 \ 400 \ 000 \ m = 2 \ 415 \ 256 \ 432.08 \ m$

The moon covers this distance in $2\ 358\ 720\ s$. Therefore, we can calculate the magnitude of the moon's velocity.

$$v = \frac{\Delta s}{\Delta t} = \frac{2\ 415\ 432.08\ \mathrm{m}}{2\ 358\ 720\ \mathrm{s}} = 1\ 023.97\ \mathrm{m.s}^{-1}$$

Now we can calculate the magnitude of the moon's momentum. $p = m \cdot v = 7.35 \times 10^{22} \text{ kg} \times 1.023.97 \text{ m.s}^{-1} = 7.53 \times 10^{25} \text{ kg.m.s}^{-1}$

Step 3: Write the final answer

The magnitude of the moon's momentum is 7.53×10^{25} kg.m.s⁻¹.



Exercise 1.1



- 1. The fastest recorded delivery for a cricket ball is 161.3 km/h, bowled by Shoaib Akhtar of Pakistan during a match against England in the 2003 Cricket World Cup, held in South Africa. Calculate the ball's momentum if it had a mass of 160 g.
- 2. The fastest tennis service by a man is 246.2 km/h by Andy Roddick of the United States of America during a match in London in 2004. Calculate the ball's momentum if it had a mass of 58 g.
- 3. Would you rather face the bowling of Shoaib Akhtar or the serve of Andy Roddick? Explain.

The <u>full solutions</u> can be found at the end of the unit.

A change in momentum

When a cricket ball hits you, it slows down. If it slows down, then its momentum changes. Whenever an object hits another object there is a change in momentum, often because of a change in the velocity of the object but, sometimes, also because of a change in the mass of the object.

It stands to reason that the change in an object's momentum is calculated as its final momentum minus its initial momentum. That is, after all what 'change' means. Therefore, $\Delta p = p_f - p_i$.

We can write this in terms of mass and velocity as $\Delta p = m_f \cdot v_f - m_i \cdot v_i$. If mass is constant (which it almost always is) we can simplify this as follows:

$$\Delta p = m_f \cdot v_f - m_i \cdot v_i$$
 $\therefore \Delta p = m \cdot v_f - m \cdot v_i$

 $\therefore \Delta p = m \cdot \Delta v$

Because momentum is a vector quantity, we can illustrate the change in momentum using the techniques of vector addition. We will deal with cases in only one dimension.

An object bouncing off a wall or the floor

Figure 1 shows the case of a ball bouncing off a wall. The total change in the object's momentum is given by the vector drawn from the end of the initial momentum vector to the end of the final momentum vector.



Figure 1: The change in momentum of an object bouncing off a wall

Figure 2 shows a similar situation of a ball bouncing off the floor. Note, however, that to keep things simple, we will ignore the effects of gravity.





Example 1.3

A netball ball with a mass of 425 g is thrown against a wall perpendicularly with a velocity of 12 m.s^{-1} . If it rebounds at a velocity of 9 m.s^{-1} , calculate its change in momentum.

Solution

Step 1: Identify the information that has been given

- \cdot The mass of the netball: $425~\mathrm{g} = 0.425~\mathrm{kg}$
- + The initial velocity of the netball: 12 m.s^{-1}
- $\cdot~$ The final velocity of the netball: _9 $_{m.s}\text{-}^1$ (We will define towards the wall as the positive direction)

We need to calculate the change in the momentum of the netball.

Step 2: Do the calculation

$$egin{aligned} \Delta p &= m \cdot \Delta v \ &= m(v_f - v_i) \ &= 0.425 \, \, \mathrm{kg}(-9 \, \mathrm{m.s^{-1}} - 12 \, \mathrm{m.s^{-1}}) \ &= 0.425 \, \, \mathrm{kg} imes - 21 \, \mathrm{m.s^{-1}} \ &= -8.925 \, \, \mathrm{kg.m.s^{-1}} \end{aligned}$$

Step 3: Write the final answer

The change in the netball's momentum is -8.925 kg.m.s⁻¹ away from the wall.

If we calculate the momentum of the netball in example 1.3 before and after it hits the wall, we get the following answers:

 $p_i = m \cdot v_i = 0.425 \text{ kg} \times 12 \text{ m.s}^{-1} = 5.1 \text{ kg.m.s}^{-1}$ towards the wall $p_f = m \cdot v_f = 0.425 \text{ kg} \times -9 \text{ m.s}^{-1} = -3.825 \text{ kg.m.s}^{-1}$ or $3.825 \text{ kg.m.s}^{-1}$ away from the wall

It may seem strange that the change in momentum is greater than either the initial or final momentum but, if you think about it, this makes sense. The netball had 5.1 kg.m.s^{-1} of momentum towards the wall to start with. It then had $3.825 \text{ kg.m.s}^{-1}$ of momentum away from the wall. Its change in momentum had to be enough to completely reverse the initial moment AND give it some momentum in the reverse direction.

An object continues but more slowly

Sometimes, when an object collides with something it does not bounce back but continues in its original direction, only more slowly. A bullet passing through a target is an example. Figure 3 shows such a case.



Figure 3: The change in momentum of an object travelling more slowly after a collision



A 40 g bullet is fired at 595 m.s^{-1} . It strikes a stationary cement block perpendicular to its surface. The bullet goes through the block and comes out on the other side at 210 m.s^{-1} . Calculate the change in momentum of the bullet.

Solution

Step 1: Identify the information that has been given

- \cdot The mass of the bullet: $40~\mathrm{g} = 0.04~\mathrm{kg}$
- $\cdot~$ The initial velocity of the bullet: $595~m.s^{-1}$
- $\cdot~$ The final velocity of the bullet: $210~m.s^{\text{-}1}$
- $\cdot\;$ Notice that both velocities are in the same direction so they both have the same sign.

We need to calculate the change in the momentum of the bullet. We will define the bullet's initial flight towards the block as the positive direction.

Step 2: Do the calculation

 $egin{aligned} \Delta p &= m \cdot \Delta v \ &= m(v_f - v_i) \ &= 0.04 \, \, \mathrm{kg} imes (210 \, \, \mathrm{m.s^{-1}} - 595 \, \, \mathrm{m.s^{-1}}) \ &= 0.04 \, \, \mathrm{kg} imes igg(-385 \, \, \mathrm{m.s^{-1}} igg) \end{aligned}$

 $= -15.4~\mathrm{kg.m.s^{-1}}$

Step 3: Write the final answer

The change in the bullet's momentum is 15.4 kg.m.s^{-1} in the opposite direction to the bullet's flight.

It might seem strange that even though the bullet continued travelling in the same direction, we still get a change in momentum in the opposite direction. However, after hitting the block, the bullet was travelling more slowly (i.e. with less momentum), hence the change of momentum is clearly negative.

An object continues but faster

Sometimes objects that are already moving are given a push to move even faster. Think of a soccer ball travelling towards the goal when a player comes up behind it and gives it another kick. Figure 4 illustrates this situation.



Figure 4: The change in momentum of an object travelling faster after a collision



 $egin{aligned} \Delta p &= m \cdot \Delta v \ &= m(v_f - v_i) \ &= 0.024 \,\, \mathrm{kg} imes (32 \,\, \mathrm{m.s^{-1}} - 1.2 \,\, \mathrm{m.s^{-1}}) \ &= 0.024 \,\, \mathrm{kg} imes 30.8 \,\, \mathrm{m.s^{-1}} \ &= 0.7392 \,\, \mathrm{kg.m.s^{-1}} \end{aligned}$

Step 3: Write the final answer

The change in the ball's momentum is $0.74~{
m kg.m.s^{-1}}$ towards the front wall.

In this case, the ball's momentum after being hit was greater than before it was hit. Therefore, the change in momentum is positive and in the direction of the front wall.



Impulse

Why do cars have airbags? What do they do? How do they improve safety? To answer these questions, we need to step back for a moment.

We know that an object in motion has momentum. We also know from experience that when a moving object (one with momentum) hits us, we feel it. It applies a force. However, Newton's third law of motion reminds us that every action has an equal and opposite reaction. In other words, if object A applies a force on object B, then object B applies an equal but opposite force on object A. We will learn more about this in the next unit.

Think of a hard ball like a cricket ball. As it hits your hand, it has momentum and applies a force to your hand. Your hand also applies a force to the ball which causes it to slow down, stop and lose all momentum. Hold this thought.

Do you remember Newton's second law of motion? It says that the force applied to an object is equal to its mass multiplied by its acceleration.

Note

If you have an internet connection, take the time to watch the video called Newton's 2nd Law of Motion. Here astronaut Randy Bresnik demonstrates what this means on the International Space Station. Newton's 2nd Law of Motion (Duration: 02.39)



It must be said that while $F_{nett} = m \cdot a$ is correct and an accurate description of Newton's second law, it is not quite how Newton described it. He described it in terms of momentum.

$$f_{nett} = m \cdot a$$

But $a = \frac{\Delta v}{t} = \frac{v - u}{t}$ or $\frac{v_f - v_i}{t}$
So $F_{nett} = m \cdot \frac{(v_f - v_i)}{t}$
But $m(v_f - v_i) = \Delta p$
Therefore $F_{nett} = \frac{\Delta p}{t}$

Note

Remember that we can refer to final and initial velocity as v and u or v_f and v_i respectively.

We can rewrite this as $\Delta p = F_{nett} \cdot t$. We call $F_{nett} \cdot t$ **impulse**. This means that $\Delta p = F_{nett} \cdot t =$ impulse.



But why do we care about impulse? If we need to change the momentum of a moving object, we can either apply a larger force for a shorter time or a smaller force for a longer time. In the case of catching a cricket ball, this matters.

Cricket players are taught to hold their hands out to the ball and gradually bring the ball into their body. They do this to increase the time over which their hands apply the force needed to reduce the ball's momentum to zero. The longer they can make this time, the less force they need to apply to the ball (and hence the ball will apply to their hands) and the less it will hurt. Cricketers call this catching the ball with 'soft hands'.

If you have an internet connection watch the video called How to Catch the Ball for a demonstration. How to Catch the Ball (Duration: 01.36)

The same is true if you ever need to catch an egg. You want to reduce the force needed to change the egg's momentum to zero, so you want to use 'soft hands' applying that force over the longest time possible.

We can now answer our initial question about air bags. Air bags reduce the effect of the force experienced by a person during an accident. If the person is stopped by a collision with the windshield, the time of the impact would be short and so a large force would need to be exerted. If instead of hitting the windshield, the person hits an air bag, then the time of the impact is increased which results in a decrease in the force.

For a constant change in momentum, like bringing a person inside a car to rest, $F_{nett} \cdot \Delta t$ is constant. Therefore, if t is increased, the force on the body is decreased.

Example 1.6

Waste pickers are pulling a trolley of mass 350 kg. They pull it with a resultant 15 N force. How long does it take for them to increase the trolley's velocity from 3 km/h to 5 km/h in the same direction?

Solution

Step 1: Identify the information that has been given

- \cdot The mass of the trolley: $350~{
 m kg}$
- + The initial velocity of the trolley: 3 $\,\mathrm{km/h} = 0.833\,\,\mathrm{m.s^{-1}}$
- + The final velocity of the trolley: 5 $\,\rm km/h = 1.389\,\,m.s^{-1}$
- + The resultant force: $15~\mathrm{N}$

We need to calculate the length of time the resultant force must be applied to accelerate the trolley from 3 km/h to 5 km/h. We will define the initial direction of the trolley as the positive direction.

Step 2: Do the calculation

Note

$$egin{aligned} F_{nett} &= rac{\Delta p}{\Delta t} \ dots &F_{nett} &= rac{m \cdot (v_f - v_i)}{\Delta t} \ dots & \Delta t &= rac{m \cdot (v_f - v_i)}{F_{nett}} \ &= rac{350 ext{ kg} imes (1.389 ext{ m.s}^{-1} - 0.833 ext{ m.s}^{-1})}{15 ext{ N}} \ &= rac{350 ext{ kg} imes 0.556 ext{ m.s}^{-1}}{15 ext{ N}} \ &= 12.97 ext{ s} \end{aligned}$$

Step 3: Write the final answer

It takes 12.97 s for the waste pickers to accelerate the trolley from 3 km/h to 5 km/h.

Example 1.7

A goalkeeper catches a soccer ball travelling at 58 km/h in 1.2 s. If the ball has a mass of 435 g, what force does he apply to the ball?

Solution

Step 1: Identify the information that has been given

- \cdot The mass of the ball: 435 g = 0.435 kg
- + The initial velocity of the ball: 58 $\rm km/h = 16.11 \ m.s^{-1}$
- $\cdot~$ The final velocity of the ball: $0~m.s^{\text{-}1}$
- + The time over which the force is applied: 1.2 $\,{\rm s}$

We need to calculate the resultant force that must be applied to catch (i.e. stop) the ball. We will define the initial direction of the ball as the positive direction.

Step 2: Do the calculation

$$egin{aligned} F_{nett} &= rac{\Delta p}{\Delta t} \ dots &: F_{nett} = rac{m \cdot (v_f - v_i)}{\Delta t} \ &= rac{0.435 \ ext{kg} imes (0 \ ext{m.s}^{-1} - 16.11 \ ext{m.s}^{-1})}{1.2 \ ext{s}} \ &= rac{0.435 \ ext{kg} imes (-16.11 \ ext{m.s}^{-1})}{1.2 \ ext{s}} \ &= -5.84 \ ext{N} \end{aligned}$$

Step 3: Write the final answer

The goalkeeper will need to apply a force of 5.84 N in the opposite direction to the ball's initial motion to catch it.

We can plot the force applied to an object against time to produce a force-time graph. These are similar

to the displacement-time, velocity-time and acceleration-time graphs of motion we learnt about in <u>level 3</u> <u>subject outcome 2.1 unit 2</u> and <u>3</u>. Figures 5 shows two examples of these graphs.



Figure 5: An example force-time graphs

What do you think the area under a force-time graph represents?

If we calculated the shaded area under either of the graphs in figure 5, the units would be N.s. In other words, the shaded areas represent impulse (impulse = $F_{nett} \cdot \Delta t$).







2. We need to calculate the impulse of the interval 3 s to 6 s. In other words, we need to calculate the area under the graph from t = 3 s to t = 6 s.







Exercise 1.3

- 1. A radio-controlled car (car A) of mass 1.3 kg is moving clockwise around a track at 12 m.s^{-1} . It collides head-on with another radio-controlled car (car B) of mass 1.12 kg moving at a velocity of 1.04 m.s^{-1} in the opposite direction. After the collision car A moves with a velocity of 1.6 m.s^{-1} in the clockwise direction and car B rebounds (travels in the opposite direction to before) at 11.03 m.s^{-1} .
 - a. Calculate the change in momentum for each toy car.
 - b. Determine the impulse for each toy car.
 - c. Determine the duration of the collision if the magnitude of the force exerted by each toy car is $16.2~{
 m N}$

- 2. A bullet of mass 22 g strikes a target at $_{330 m.s}$ -1 and exits at $_{210 m.s}$ -1. The tip of the bullet takes 0.00013 s to pass through the target. Determine:
 - a. The change of momentum of the bullet.
 - b. The impulse of the bullet.
 - c. The magnitude of the force experienced by the bullet.
- 3. A piece of modelling clay with a mass of $400~{
 m g}$ is thrown against a wall with a velocity of $9~{
 m m.s}^{-1}$. It sticks to the wall. Calculate:
 - a. The impulse of the clay on the wall.
 - b. The force exerted by the clay on the wall if it is in contact with the wall for $0.25~{
 m s}$ before it comes to rest.

The <u>full solutions</u> can be found at the end of the unit.

Summary

In this unit you have learnt the following:

- That momentum is mass in motion. Every moving object has momentum.
- We define momentum (p) as the mass of an object multiplied by its velocity. Therefore, $p = m \cdot v$.
- Momentum is a vector.
- Momentum has units of $kg.m.s^{-1}$.
- The change in momentum is $\Delta p = m \cdot \Delta v$.
- $\cdot \ \ \Delta p = F_{nett} \cdot t = ext{impulse}$
- The area under a force-time graph represents impulse.

Unit 1: Assessment

Suggested time to complete: 35 minutes

Question 1 adapted from Everything Science Grade 12 Worked example 16 page 55

- 1. A cricket ball weighing 156 g is moving at 54 km/h towards a batsman. It is hit by the batsman back towards the bowler at 36 km/h. Calculate:
 - a. The ball's impulse.
 - b. The average force exerted by the bat if the ball is in contact with the bat for $0.13~{
 m s}$.
- 2. A squash player hits the ball (mass 24 g) and it strikes her opponent at 105 $\rm km/h$ on the leg. It rebounds at 43 $\rm km/h$.
 - a. Calculate the change in the momentum of the ball.
 - b. What is the impulse of the ball on the player?
 - c. Calculate the magnitude of the force exerted by the ball on the player if the collision takes $0.017~{
 m s}$.
- 3. Modern cars are deliberately designed with a front crumple zone an area designed to deliberately crumple in a head on collision Why is this a good idea?

The <u>full solutions</u> can be found at the end of the unit.

Unit 1: Solutions

Exercise 1.1

l. m = 160 g = 0.160 kg $v = 161.3 \text{ km/h} = 44.81 \text{ m.s}^{-1}$

Define the direction in which the ball is bowled as positive. $p = m \cdot v = 0160 \text{ kg} \times 44.81 \text{ m.s}^{-1} = 7.17 \text{ kg.m.s}^{-1}$

The ball's momentum is 7.17 kg.m.s^{-1} in the positive direction.

2. m = 58 g = 0.058 kg $v = 246.2 \text{ km/h} = 68.39 \text{ m.s}^{-1}$

> Define the direction in which the ball is served as positive. $p = m \cdot v = 0.058 \text{ kg} \times 68.39 \text{ m.s}^{-1} = 3.97 \text{ kg.m.s}^{-1}$

The ball's momentum is 3.97 kg.m.s^{-1} in the positive direction.

3. The cricket ball bowled by Shoaib Akhtar had more momentum than the tennis ball served by Andy Roddick. Therefore, it would be better to face the serve from Andy Roddick because, if the ball did hit you, it would hurt less.

Back to Exercise 1.1

Exercise 1.2

1. $m = 590 \, \mathrm{g} = 0.590 \, \mathrm{kg}$

 $v_i=7\,\,{
m m.s}^{-1}$ towards the ground

 $v_i = 4.2\,\,{
m m.s^{-1}}$ away from the ground

Define towards the ground as positive. $\Delta p = m(v_f - v_i) = 0.590~{\rm kg} \times (-4.2~{\rm m.s^{-1}} - 7~{\rm m.s^{-1}}) = -6.61~{\rm kg.m.s^{-1}}$

The ball's change in momentum is $6.61~{
m kg.m.s^{-1}}$ away from the ground.

2. $m=2.2~{
m kg}$ $\Delta p=-5.73~{
m kg.m.s}^{-1}$

 $v_f = 13 \,\,\mathrm{m.s}^{-1}$

Define the brick's direction towards and through the glass as positive. Note that the change in momentum is negative. The brick will travel more slowly after hitting the glass. Therefore, its momentum will be less. Because it will still be travelling in the same (positive) direction, this change (or loss) of momentum will be negative.

$$egin{aligned} \Delta p &= m(v_f - v_i) \ dots v_f - v_i &= rac{\Delta p}{m} \ dots v_i &= v_f - rac{\Delta p}{m} \ &= 13 \,\,\mathrm{m.s^{-1}} - rac{(-5.73 \,\,\mathrm{kg.m.s^{-1}})}{2.2 \,\,\mathrm{kg}} \ &= 15.60 \,\,\mathrm{m.s^{-1}} \end{aligned}$$

The brick's velocity just before hitting the glass was $15.60~{
m m.s}^{-1}$ in the positive direction.

Back to Exercise 1.2

Exercise 1.3

1.

a. Car A: m = 1.3 kg $v_i = 12 \text{ m.s}^{-1} \text{ clockwise}$ $v_f = 1.6 \text{ m.s}^{-1} \text{ clockwise}$

Define clockwise as positive. $\Delta p = m(v_f - v_i) = 1.3 \ {
m kg} imes (1.6 \ {
m m.s}^{-1} - 12 \ {
m m.s}^{-1}) = -13.52 \ {
m kg.m.s}^{-1}$

The change in momentum of car A is $13.52 \text{ kg.m.s}^{-1}$ in the anti-clockwise direction.

Car B: m = 1.12 kg $v_i = 1.04 \text{ m.s}^{-1}$ anti-clockwise $v_f = 11.03 \text{ m.s}^{-1}$ m.s-1 clockwise

Define clockwise as positive. $\Delta p = m(v_f - v_i) = 1.12 \text{ kg} imes (11.03 \text{ m.s}^{-1} - (-1.04) \text{ m.s}^{-1}) = 13.52 \text{ kg.m.s}^{-1}$

The change in momentum of car B is $13.52 \text{ kg.m.s}^{-1}$ in the clockwise direction.

b. impulse = Δp

 $\label{eq:Car_A} \begin{array}{l} \mathrm{impulse_{Car\ A}} = -13.52 \ \mathrm{kg.m.s^{-1}} \ \text{or} \ 13.52 \ \mathrm{kg.m.s^{-1}} \ \text{in the anti-clockwise direction} \\ \mathrm{impulse_{Car\ B}} = 13.52 \ \mathrm{kg.m.s^{-1}} \ \text{or} \ 13.52 \ \mathrm{kg.m.s^{-1}} \ \text{in the clockwise direction} \\ \end{array}$

c. As the magnitude of the change in momentum is the same for both cars, we can use the change in momentum for either car in the calculation. We are calculating time and so direction is not required.

 $F_{nett} = 16.2 \,\, \mathrm{N} \ \Delta p = 13.52 \,\, \mathrm{kg.m.s}^{-1}$

$$egin{aligned} \mathrm{impulse} &= F_{nett} \cdot \Delta t \\ \therefore \Delta t &= rac{\Delta p}{F_{nett}} = rac{13.52 \ \mathrm{kg.m.s^{-1}}}{16.2 \ \mathrm{N}} = 0.83 \ \mathrm{s} \end{aligned}$$

The duration of the collision was $0.83 \, \mathrm{s}$.

2.

a. m = 22 g = 0.022 kg $v_i = 330 \text{ m.s}^{-1}$ $v_f = 210 \text{ m.s}^{-1}$

Define the bullet's direction towards and through the target as positive. $\Delta p = m(v_f - v_i) = 0.022 \text{ kg} \times (210 \text{ m.s}^{-1} - 330 \text{ m.s}^{-1}) = -2.64 \text{ kg.m.s}^{-1}$

The change in the momentum of the bullet is 2.64 kg.m.s^{-1} in the negative direction.

b. impulse = $\Delta p = -2.64$ kg.m.s⁻¹ Impulse is 2.64 kg.m.s⁻¹ in the negative direction.
C. $\Delta t = 0.00013~{
m s}$ $\Delta p = -2.64~{
m kg.m.s}^{-1}$

$$egin{aligned} ext{impulse} &= F_{nett} \cdot \Delta t \ dots &: F_{nett} = rac{\Delta p}{\Delta t} = rac{-2.64 ext{ kg.m.s}^{-1}}{0.00013 ext{ s}} = -20 ext{ 307.70 N} \end{aligned}$$

The bullet experienced a force of 20 307.70 N in the opposite direction to its motion.

3.

a. m = 400 g = 0.4 kg $v_i = 9 \text{ m.s}^{-1}$ $v_f = 0 \text{ m.s}^{-1}$

Define towards the wall as positive. $\Delta p = m(v_f - v_i) = 0.4 \ {
m kg} imes (0 \ {
m m.s}^{-1} - 9 \ {
m m.s}^{-1}) = -3.6 \ {
m kg.m.s}^{-1}$

The impulse of the clay is 3.6 kg.m.s^{-1} in the opposite direction to its initial motion.

b.
$$\Delta t = 0.25 ext{ s}$$

 $\Delta p = -3.6 ext{ kg.m.s}^{-1}$

$$egin{aligned} & ext{impulse} = F_{nett} \cdot \Delta t \ & ext{.} \ F_{nett} = rac{\Delta p}{\Delta t} = rac{-3.6 ext{ kg.m.s}^{-1}}{0.25 ext{ s}} = -14.4 ext{ N} \end{aligned}$$

The force exerted by the wall on the clay is 14.4 N in the opposite direction to the clay's initial motion. Therefore, the force exerted by the clay on the wall is 14.4 N in the direction of the clay's original motion because the force exerted by the wall on the clay will be equal in magnitude but opposite in direction to the force exerted by the clay on the wall.

Back to Exercise 1.3

Unit 1: Assessment

Question 1 adapted from Everything Science Grade 12 Worked example 16 page 55

1.

a. m = 156 g = 0.156 kg $v_i = 54 \text{ km/h} = 15 \text{ m.s}^{-1}$ $v_f = -36 \text{ km/h} = -10 \text{ m.s}^{-1}$

Define from bowler to batsman as positive. $\Delta p = m(v_f - v_i) = 0.156~{
m kg} imes (-10~{
m m.s}^{-1} - 15~{
m m.s}^{-1}) = -3.9~{
m kg.m.s}^{-1}$

The ball's impulse (or change in momentum) is $3.9~{
m kg.m.s^{-1}}$ from batsman to bowler.

b. $\Delta t = 0.13 ext{ s}$ $\Delta p = -3.9 ext{ kg.m.s}^{-1}$

 $egin{aligned} &\mathrm{impulse} = F_{nett} \cdot \Delta t \ &\therefore F_{nett} = rac{\Delta p}{\Delta t} = rac{-3.9 \mathrm{~kg.m.s^{-1}}}{0.13 \mathrm{~s}} = -30 \mathrm{~N} \end{aligned}$

The average force exerted by the bat is $30\,\,\mathrm{N}$ in the direction batsman to bowler.

2.

a. m = 24 g = 0.024 kg

 $v_i = 105 \, {
m km/h} = 29.17 \, {
m m.s}^{-1}
onumber v_f = -43 \, {
m km/h} = -11.94 \, {
m m.s}^{-1}$

Define from racquet towards player who gets hit as positive. $\Delta p = m(v_f - v_i) = 0.024 \text{ kg} \times (-11.94 \text{ m.s}^{-1} - 29.17 \text{ m.s}^{-1}) = -0.99 \text{ kg.m.s}^{-1}$

The change in the ball's momentum is 0.99 kg.m.s^{-1} from the hit player to the racquet.

- b. impulse = $\Delta p = -0.99$ kg.m.s⁻¹ in the direction from the hit player to the racquet.
- C. $\Delta t = 0.017~{
 m s}$ $\Delta p = -0.99~{
 m kg.m.s}^{-1}$

$$egin{aligned} \operatorname{impulse} &= F_{nett} \cdot \Delta t \ &\therefore F_{nett} &= rac{\Delta p}{\Delta t} &= rac{-0.99 \, \mathrm{kg.m.s^{-1}}}{0.017 \; \mathrm{s}} &= -58.24 \; \mathrm{N} \end{aligned}$$

The force exerted by the player's leg on the ball is 58.24 N in the opposite direction to the ball's initial motion. Therefore, the force exerted by the ball on the player's leg is 58.24 N in the direction of the ball's original motion (from racquet to hit player) because the force exerted by the leg on the ball will be equal in magnitude but opposite in direction to the force exerted by the ball on the leg.

3. A crumple zone will reduce the impulse during a collision by extending the time over which the vehicle's velocity reduces and hence the time over which its momentum changes. Because impulse $= \Delta p = F_{nett} \cdot \Delta t$, if t increases for a constant Δp then F_{nett} will decrease and less force will be applied to the passengers in the car.

Back to Unit 1: Assessment

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Unit 2: Collisions

DYLAN BUSA



By the end of this unit you will be able to:

- Define Newton's third law and identify examples of interacting objects that exert equal forces on each other.
- Define the law of conservation of momentum.
- Distinguish between elastic and inelastic collisions.
- Apply the law of conservation of momentum in calculations on different types of collisions.

What you should know

Before you start this unit, make sure you can:

- Define and calculate momentum. Refer to <u>unit 1</u> if you need help with this.
- · Calculate the change of momentum of an object.
- Define impulse as $F_{nett} \cdot t$ or Δp . Refer to <u>unit 1</u> if you need help with this.
- Understand kinetic energy. Refer back to level 2 subject outcome 2.3 if you need help with this.

Introduction

In the previous unit, we learnt how Isaac Newton viewed his second law of motion (normally expressed as

$$F_{nett}=m\cdot a$$
) in terms of momentum, in other words that $F_{nett}=rac{\Delta p}{\Delta t}$ or $\Delta p=F_{nett}\cdot \Delta t.$

We defined $\Delta p = F_{nett} \cdot \Delta t$ as impulse and looked at several situations where we calculated either the force or the time necessary to effect a specific change in momentum.

In many of these cases, we had two objects colliding with each other where the force exerted by object A on object B is equal and opposite to the force exerted by object B on object A. This is a simple statement of Newton's third law of motion. But this has important consequences for momentum as we will see.

Newton's third law of motion

You have probably heard the expression that 'every action has an equal and opposite reaction'. This is a statement of Newton's third law of motion. You have experienced this law of motion every time you hit your head on something. Your head applies a force to the object but the object applies an equal but opposite force to your head and it hurts!

Note

If you have an internet connection, watch the video called Newton's Third Law of Motion for an excellent demonstration of the third law.

Newton's Third Law of Motion (Duration: 03.07)



Think about the case where two billiard balls collide (see figure 1).



We can express the third law as $F_{3 \text{ ball}} = -F_{8 \text{ ball}}$. But we also know from the previous unit that $F_{nett} = \frac{\Delta p}{\Delta t}$.

Therefore, we can say that:

$$egin{array}{ll} \displaystylerac{\Delta p_{3\ ext{ball}}}{\Delta t} &= -rac{\Delta p_{8\ ext{ball}}}{\Delta t} \ \displaystylerac{\Delta p_{3\ ext{ball}}}{\Delta t} \ \displaystylerac{\Delta p_{3\ ext{ball}}}{\Delta t} &= -\Delta p_{8\ ext{ball}} \ \displaystylerac{\Delta p_{3\ ext{ball}}}{\Delta t} &= -\Delta p_{8\ ext{ball}} \ \displaystylerac{\Delta p_{3\ ext{ball}}}{\Delta t} &= 0 \end{array}$$

This means that if you add up all the changes in momentum for an isolated system the net result will be zero. If we add up all the momenta in the system, the total momentum won't change because the net change is zero.

The conservation of momentum

Another way of saying this is that, in a closed system, momentum is conserved or the total momentum before a collision (or explosion) is equal to the total momentum after a collision (or explosion). We can express this as $P_{Ti} = P_{Tf}$.

The total momentum of the system before a collision is $m_1 \cdot v_{i1} + m_2 \cdot v_{i2}$ and the total momentum of the system after the collision is $m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$ (see figure 2).



Figure 2: The total momentum is conserved

This means that, where there are two colliding objects $m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$. The momentum of the isolated system is **always** conserved.

Note
In an isolated system with two colliding objects $p_T = m_1 \cdot v_1 + m_2 \cdot v_2.$
But $p_i=p_f$. Therefore $m_1\cdot v_{i1}+m_2\cdot v_{i2}=m_1\cdot v_{f1}+m_2\cdot v_{f2}.$
The momentum of an isolated system is always conserved.
Let's explore this further in the next two examples.



In the situation illustrated in figure 1, the mass of each ball is 300 g. The 3 ball is initially moving at 2 m.s^{-1} to the right while the 8 ball is stationary.

- 1. What is the total momentum of the system before the collision?
- 2. If, after the collision, the 8 ball moves to the right at 2 m.s^{-1} , what is the velocity of the 3 ball?

Solutions

1. **Step 1**: Identify the information that has been given 8 ball:

 $v_{i8}=0 \mathrm{~m.s^{-1}}$. Therefore, $p_{i8}=0 \mathrm{~kg.m.s^{-1}}$.

3 ball:

 $m_3 = 300 \,\mathrm{g} = 0.3 \,\mathrm{kg}$ $v_{i3} = 2 \,\mathrm{m.s}^{-1}$ to the right We define to the right as positive.

Step 2: Do the calculation

 $p_{i3} = m_3 \times v_{i3} = 0.3 \text{ kg} \times 2 \text{ m.s}^{-1} = 0.6 \text{ kg.m.s}^{-1}$ The total momentum of the system will be the momentum of the 3 ball plus the momentum of the 8 ball.

 $p_{Ti} = p_{i3} + p_{i8} = 0.6 \, \, {
m kg.m.s^{-1}} + 0 \, \, {
m kg.m.s^{-1}} = 0.6 \, \, {
m kg.m.s^{-1}}$

Step 3: Write the final answer The total momentum of the system before the collision is 0.6 kg.m.s^{-1} to the right.

2. We know that the change in momentum of this closed or isolated system is zero. Therefore, the total momentum of the system before the collision will be equal to the total momentum of the system after the collision.

Step 1: Identify the information that has been given $p_{Ti} = p_{Tf} = 0.6 \text{ kg.m.s}^{-1}$ $v_{f8} = 2 \text{ m.s}^{-1}$ to the right $m_3 = 300 \text{ g} = 0.3 \text{ kg}$

Step 2: Do the calculation $p_{f8} = 0.3 \ \mathrm{kg} \times 2 \ \mathrm{m.s}^{-1} = 0.6 \ \mathrm{kg.m.s}^{-1}$

 $p_{f8} = 0.6 \text{ kg} \times 2 \text{ m.s}^{-1} = 0.6 \text{ kg.m.s}^{-1}$ But $p_{Tf} = p_{f8} + p_{f3} = 0.6 \text{ kg.m.s}^{-1}$

 $\therefore p_{f3} = p_{tf} - p_{f8}$ $\therefore p_{f3} = 0.6 ext{ kg.m.s}^{-1} - 0.6 ext{ kg.m.s}^{-1} = 0 ext{ kg.m.s}^{-1}$

Step 3: Write the final answer The 3 ball will move at $_0\,\,{\rm m.s^{-1}}$ immediately after the collision. It will be stationary.



Example adapted from Everything Science Grade 12 Worked example 9 page 42

A jet flies at a speed of $_{275 m.s}$ -1 relative to the ground. The pilot fires a missile forward off a mounting at a speed of $_{700 m.s}$ -1 relative to the ground. The respective masses of the jet and the missile are 5 000 kg and 50 kg.



Treating the system as an isolated system, calculate the new speed of the jet immediately after the missile was fired.

Solution

Step 1: Identify the information that has been given Let the plane be object 1 and the missile object 2. Define the initial direction of the plane as positive.

 $m_1 = 5\ 000 \text{ kg}$ $v_{i1} = 275 \text{ m.s}^{-1}$ $m_2 = 50 \text{ kg}$ $v_{i2} = v_{i1} = 275 \text{ m.s}^{-1}$ (the missile and plane are travelling at the same speed) $v_{f2} = 700 \text{ m.s}^{-1}$ We need to find the final speed of the plane (v_{f1}) and we can use the conservation of momentum because we can treat it as an isolated system.

Step 2: Do the calculation

$$p_i = p_f$$

 $\therefore m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{fi} + m_2 \cdot v_{f2}$
But $v_{i2} = v_{i1}$
 $\therefore (m_1 + m_2) \cdot v_{i1} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$
 $\therefore m_1 \cdot v_{f1} = (m_1 + m_2) \cdot v_{i1} - m_2 \cdot v_{f2}$
 $\therefore v_{f1} = \frac{(m_1 + m_2) \cdot v_{i1} - m_2 \cdot v_{f2}}{m_1}$
 $\therefore v_{f1} = \frac{(5 \ 000 \ \text{kg} + 50 \ \text{kg}) \times 275 \ \text{m.s}^{-1} - (50 \ \text{kg} \times 700 \ \text{m.s}^{-1})}{5 \ 000 \ \text{kg}}$
 $= 270.75 \ m. s^{-1}$

Step 3: Write the final answer

The plane's velocity after firing the missile is 270.75 m.s^{-1} in its initial direction.



Exercise 2.1

- 1. A 70 ton train is travelling north at a velocity of 40 km/h when it is approached by a 50 ton train travelling south at a velocity of 55 km/h in the opposite direction. Calculate the total momentum of the system.
- 2. A bullet of mass 55 g travelling horizontally at 540 m.s^{-1} hits a stationary target of mass 3.5 kg resting on a smooth horizontal surface. The bullet goes through the target and comes out on the other side at 225 m.s^{-1} . Calculate the velocity of the target after the bullet has come out the other side. Assume that the target experiences no friction in its motion.

The <u>full solutions</u> can be found at the end of the unit.

Elastic collisions

Have a look at <u>example 2.1</u> again. Here we had two billiard balls colliding. We were told that each ball has a mass of 300 g and that initially the 3 ball rolled to the right at 2 m.s^{-1} while the 8 ball was stationary. After the collision, the 8 ball moved to the right with a velocity of 2 m.s^{-1} and we calculated, because of the conservation of momentum, that the 3 ball would move to the right with a velocity of 0 m.s^{-1} (see figure 3).



Figure 3: The collision of the 3 ball and 8 ball

We know that momentum in this isolated system is conserved. But what else is conserved? Have a look at figure 3 to see if you can see what this is.

The mass of the balls is the same. We also know that billiard balls are made to transfer kinetic energy very effectively from one ball to the other. Can you see that the total kinetic energy before the collision is the same as the total kinetic energy after the collision?

Before the collision:

$$egin{aligned} & KE_{i3} = rac{1}{2} imes m_3 imes v_{i3}{}^2 = rac{1}{2} imes 0.3 \,\, ext{kg} imes ig(2 \,\, ext{m.s}{}^{-1}ig)^2 = 0.6 \,\, ext{J} \ & KE_{i8} = rac{1}{2} imes m_8 imes v_{i8}{}^2 = rac{1}{2} imes 0.3 \,\, ext{kg} imes ig(0 \,\, ext{m.s}{}^{-1}ig)^2 = 0 \,\, ext{J} \end{aligned}$$

$$KE_{Ti} = 0.6$$
 J

After the collision:

$$egin{aligned} & KE_{f3} = rac{1}{2} imes m_3 imes v_{f3}{}^2 = rac{1}{2} imes 0.3 \,\, ext{kg} imes ig(0 \,\, ext{m.s}{}^{-1}ig)^2 = 0 \,\, ext{J} \ & KE_{f8} = rac{1}{2} imes m_8 imes v_{f8}{}^2 = rac{1}{2} imes 0.3 \,\, ext{kg} imes ig(2 \,\, ext{m.s}{}^{-1}ig)^2 = 0.6 \,\, ext{J} \end{aligned}$$

 $KE_{Tf} = 0.6$ J

This is an example of an **elastic** collision. This is a collision where the total momentum of the isolated system is conserved AND the total kinetic energy of the isolated system is conserved (see figure 4).



Figure 4: An example of an elastic collision - the total momentum and total kinetic energy are conserved

(**D**) Take note!

In an elastic collision:

· total momentum is conserved: $p_i = p_f$ · $p_{i1} + p_{i2} = p_{i1} + p_{i2}$

$$\dots p_{i1} + p_{i2} - p_{f1} + p_{f2}$$

 $\therefore m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$

• total kinetic energy is conserved: EK - EK

$$EK_i = EK_f \ \therefore EK_{i1} + EK_{i2} = EK_{f1} + EK_{f2} \ \therefore rac{1}{2} \cdot m_1 \cdot v_{i1}^2 + rac{1}{2} \cdot m_2 \cdot v_{i2}^2 = rac{1}{2} \cdot m_1 \cdot v_{f1}^2 + rac{1}{2} \cdot m_2 \cdot v_{f2}^2$$



Example 2.3

Two marbles collide. Marble 1 has mass $30~{
m g}$ and marble 2 has mass $60~{
m g}$.

Marble 1 rolls to the right with an initial velocity of 2.5 m.s^{-1} . Marble 2 is initially at rest. After they collide, both marbles are moving.

- 1. What is the final velocity of marble 1 if marble 2 has a velocity of $1.8 \ m.s^{-1}$ to the right?
- 2. Is this an elastic collision?

Solutions

1. Step 1: Identify the information that has been given $m_1 = 30 \text{ g} = 0.03 \text{ kg}$ $v_{i1} = 2.5 \text{ m.s}^{-1}$ to the right $m_2 = 60 \text{ g} = 0.06 \text{ kg}$ $v_{i2} = 1.8 \text{ m.s}^{-1}$ $v_{f2} = 1.1 \text{ m.s}^{-1}$ Define movement to the right as positive.

Step 2: Do the calculation
Momentum conserved:

$$m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$$

But $v_{i2} = 0 \text{ m.s}^{-1}$
 $\therefore m_1 \cdot v_{i1} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$
 $\therefore m_1 \cdot v_{f1} = m_1 \cdot v_{i1} - m_2 \cdot v_{f2}$
 $\therefore v_{f1} = \frac{m_1 \cdot v_{i1} - m_1 \cdot v_{f2}}{m_1}$
 $= \frac{0.003 \text{ kg} \times 2.5 \text{ m.s}^{-1} - 0.06 \text{ kg} \times 1.1 \text{ m.s}^{-1}}{0.03 \text{ kg}}$
 $= 0.3 \text{ m.s}^{-1}$

Step 3: Write the final answer Marble 1 will have a final velocity of $0.3\ {\rm m.s}^{\text{-1}}$ to the right.

2. To determine if the collision is elastic we need to determine if kinetic energy is conserved.

$$\begin{split} EK_i &= \frac{1}{2} \cdot m_1 \cdot v_{i1}^2 + \frac{1}{2} \cdot m_2 \cdot v_{i2}^2 \\ &= \frac{1}{2} \times 0.03 \text{ kg} \times (2.5 \text{ m.s}^{-1})^2 + \frac{1}{2} \times 0.06 \text{ kg} \times (0 \text{ m.s}^{-1})^2 \\ &= 0.09375 \text{ J} \\ EK_f &= \frac{1}{2} \cdot m_1 \cdot v_{f1}^2 + \frac{1}{2} \cdot m_2 \cdot v_{f2}^2 \\ &= \frac{1}{2} \times 0.03 \text{ kg} \times (0.3 \text{ m.s}^{-1})^2 + \frac{1}{2} \times 0.06 \text{ kg} \times (1.1 \text{ m.s}^{-1})^2 \\ &= 0.00135 \text{ J} + 0.0363 \text{ J} \\ &= 0.03765 \text{ J} \\ EK_i \neq EK_f. \text{ Therefore, the collision is not elastic.} \end{split}$$

Take note!

If you are asked to prove that a collision is elastic, you need to show that the total kinetic energy before the collision is equal to the total kinetic energy after the collision.

Inelastic collisions

In reality, however, most collisions are not elastic. In fact, the only perfectly elastic collisions that occur in the real world are those between some sub-atomic particles. In all other cases, the collisions result in some loss of kinetic energy in the form of sound, heat, and the permanent deformation of the objects in question.

Figure 5 shows an example of this. In this head on collision between these two cars, it is clear that they did not bounce off each other like billiard balls! Most of the initial kinetic energy before the collision was converted to sound, heat, and the energy necessary to permanently deform the cars. The force needed to damage the cars was available because of the conservation of momentum. Remember that in all collisions, momentum is always conserved.



Figure 5: A non-elastic collision

Another example of an inelastic collision is a meteor crashing into the earth and forming an impact crater (see figure 6). In this kind of collision, the initial kinetic energy of the meteor is definitely not conserved as kinetic energy. Instead, it results in so much heat being generated, that the meteor and the earth's surface is largely melted or vapourised and a huge crater in the surface of the earth is formed.



Figure 6: An impact crater



A truck and a car are involved in a head on collision. The car is initially travelling at 65 km/h east while the truck is travelling at 30 km/h west. The truck has a mass of 3.5 tons while the car has a mass of 850 kg. After colliding inelastically, the truck and the car stick together. What is the resulting velocity of the mass of metal formed?

Solution

Step 1: Identify the information that has been given Let the truck be object 1 and the car be object 2. Define west as positive $m_1 = 3.5 \text{ tons} = 3500 \text{ kg}$ $v_{i1} = 30 \text{ km/h} = 8.33 \text{ m.s}^{-1}\text{West}$ $m_2 = 850 \text{ kg}$ $v_{i2} = 65 \text{ km/h} = 18.06 \text{ m.s}^{-1}\text{east}$ The truck and the car become one mass after the collision. Therefore, $v_{f1} = v_{f2} = v_f$. We need to calculate the velocity of the car and truck mass. Momentum is conserved. **Step 2**: Do the calculation

 $egin{aligned} &m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ dots \cdot m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = v_f \cdot (m_1 + m_2) \ dots \cdot v_f &= rac{m_1 \cdot v_{i1} + m_2 \cdot v_{i2}}{(m_1 + m_2)} \ &= rac{3\ 500\ \mathrm{kg} imes 8.33\ \mathrm{m.s^{-1}} + 850\ \mathrm{kg} imes (-18.06\ \mathrm{m.s^{-1}})}{3\ 500\ \mathrm{kg} + 850\ \mathrm{kg}} \ &= 3.17\ \mathrm{m.s^{-1}} \end{aligned}$

Step 3: Write the final answer The final velocity of the truck and car mass is $3.17\ m.s^{-1}$ west.



Take note!

If you are asked to prove that a collision is inelastic, you need to show that the total kinetic energy before the collision is not equal to the total kinetic energy after the collision.



Exercise 2.2

Question 1 adapted from Everything Science Grade 12 Exercise 2 – 4 question 1

- 1. A truck of mass 4 500 kg travelling at $_{20 m.s}$ -1 hits a car from behind. The car (mass 1 000 kg) is travelling at $_{15 m.s}$ -1. The two vehicles, now connected, carry on moving together.
 - a. Calculate the final velocity of the truck-car combination after the collision.
 - b. Determine the kinetic energy of the system before and after the collision.
 - c. Explain the difference in your answers for b.
 - d. Was this an example of an elastic or inelastic collision? Give reasons for your answer.
- 2. Two billiard balls each with a mass of 200 g collide head-on in an elastic collision. The 4 ball was travelling at a velocity of 2.5 m.s^{-1} and the 2 ball at a velocity of 1.5 m.s^{-1} . After the collision, the 4 ball travels away from the 2 ball at a velocity of 1.5 m.s^{-1} .
 - a. Calculate the velocity of the 2 ball after the collision.

b. Prove that the collision was elastic, showing all calculations.

The <u>full solutions</u> can be found at the end of the unit.

Summary

In this unit you have learnt the following:

- Newton's third law of motion states that every action has an equal and opposite reaction. In other words, if object A exerts a force on object B, then object B will exert an equal but opposite force on object A.
- Momentum is conserved. This means that the momentum of any isolated system is constant or the change in momentum of an isolated system is zero or the total momentum before a collision is equal to the total momentum after a collision.
- In elastic collisions kinetic energy is conserved or the kinetic energy before a collision is equal to the kinetic energy after a collision.
- In inelastic collisions kinetic energy is not conserved. The kinetic energy before a collision is not equal to the kinetic energy after a collision.

Unit 2: Assessment

Suggested time to complete: 15 minutes

Question 1 adapted from Everything Science Grade 12 Worked example 18 page 59

- 1. A thief is driving at 40 m.s^{-1} east to evade the police. The thief's car has a mass of $1\ 000$ kg. The thief's car collides head-on with a truck of mass $5\ 000$ kg moving at $20\ \text{m.s}^{-1}$ in the opposite direction. After the collision, the car and the truck move together. Ignore the effects of friction.
 - a. State the law of conservation of linear momentum in words.
 - b. Calculate the velocity of the thief's car immediately after the collision.
 - c. Research has shown that forces greater than 85 000 N during collisions may cause fatal injuries. The collision described above lasts for 0.5 s. Determine, by means of a calculation, whether the collision above could result in a fatal injury.

Question 2 adapted from NC(V) Physical Science Level 4 November 2015 Paper 1 question 7.2

2. The figure shows a car with a mass of 1 000 kg that moves along a horizontal road at $_{20 m.s}$ -1 west. It collides with a taxi of mass 1 500 kg moving along the same horizontal road at $_{25 m.s}$ -1 east. The two vehicles join together upon collision.



- a. Calculate the velocity of the car after collision.
- b. A learner says that the force that the car and the taxi experience during collision will be the same. Is the learner correct? Explain the answer in terms of the relevant law of physics.
- c. Calculate the impulse experienced by the taxi.
- d. Is the collision elastic or inelastic? Show by means of a calculation how you arrived at the answer.

Question 3 adapted from NC(V) Physical Science Level 4 November 2019 Paper 1 question 7

- 3. An object of mass 3 kg, moving at $_{6 m.s^{-1}}$, collides with a stationary object of mass 6 kg. The 3 kg object rebounds after the collision with a speed of $_{2 m.s^{-1}}$.
 - a. Calculate the speed of the 6 kg object after collision.
 - b. Show by means of a calculation whether the collision is elastic or inelastic.
- 4. In a railway shunting yard, a locomotive of mass 4 500 kg, travelling north at a velocity of 2.5 m.s^{-1} , collides with a stationary goods wagon of mass 7 000 kg in an attempt to couple with it. The coupling fails and instead the goods wagon moves north with a velocity of 1.5 m.s^{-1} .
 - a. Calculate the magnitude and direction of the velocity of the locomotive immediately after collision.
 - b. Name and state in words the law you used to answer the previous question.
 - c. Was the collision inelastic or not? Show your calculations.
 - d. If the coupling had been successful, what would the locomotive's velocity have been immediately after the collision?

The <u>full solutions</u> can be found at the end of the unit.

Unit 2: Solutions

Exercise 2.1

1. Let train travelling north be 1 and train travelling south be 2. $m_1 = 70 \text{ ton} = 70 000 \text{ kg}$ $v_1 = 40 \text{ km/h} = 11.11 \text{ m.s}^{-1} \text{ north}$ $m_2 = 50 \text{ ton} = 50 000 \text{ kg}$ $v_2 = 55 \text{ km/h} = 15.28 \text{ m.s}^{-1} \text{ south}$ Let north be positive. We need to calculate the total momentum. $p_T = p_1 + p_2$

```
= m_1 \cdot v_1 + m_2 \cdot v_2
= 70 000 kg × 11.11 m.s<sup>-1</sup> + 50 000 kg × (-15.28 m.s<sup>-1</sup>)
= 13 700 kg.m.s<sup>-1</sup>
The total momentum is 13 700 kg.m.s<sup>-1</sup> north.
```

2. Let the bullet be object 1 and target be object 2.

 $m_1 = 55 \text{ g} = 0.055 \text{ kg}$ $v_{i1} = 540 \text{ m.s}^{-1}$ $v_{f1} = 225 \text{ m.s}^{-1}$ $m_2 = 3.5 \text{ kg}$ $v_{i2} = 0 \text{ m.s}^{-1}$ Let the bullet's initial direction be positive.

$$egin{aligned} m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ & ext{But } v_{i2} &= 0 \ m. \ s^{-1} \ dots \cdot m_1 \cdot v_{i1} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ dots \cdot m_2 \cdot v_{f2} &= m_1 \cdot v_{i1} - m_1 \cdot v_{f1} \ dots \cdot v_{f2} &= rac{m_1 \cdot v_{i1} - m_1 \cdot v_{f1}}{m_2} \ dots \cdot v_{f2} &= rac{m_1 \cdot (v_{i1} - v_{f1})}{m_2} \ dots \cdot v_{f2} &= rac{m_1 \cdot (v_{i1} - v_{f1})}{m_2} \ &= rac{0.055 \ ext{kg} imes (540 \ ext{m.s}^{-1} - 225 \ ext{m.s}^{-1})}{3.5 \ ext{kg}} \ &= 4.95 \ ext{m.s}^{-1} \end{aligned}$$

The block's velocity will be $4.95~{
m m.s^{-1}}$ in the direction of the bullet.

Back to Exercise 2.1

Exercise 2.2

1.

a. Let the truck be object 1 and the car be object 2. Let the initial direction of the truck and car be positive. $m_1 = 4\ 500 \text{ kg}$ $v_{i1} = 20 \text{ m.s}^{-1}$ $m_2 = 1\ 000 \text{ kg}$ $v_{i2} = 15 \text{ m.s}^{-1}$ $v_{f1} = v_{f2} = v_f$ $m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2}$ $\therefore m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = v_f \cdot (m_1 + m_2)$ $\therefore v_f = \frac{m_1 \cdot v_{i1} + m_2 \cdot v_{i2}}{(m_1 + m_2)}$ $= \frac{4\ 500\ \text{kg} \times 20\ \text{m.s}^{-1} + 1\ 000\ \text{kg} \times 15\ \text{m.s}^{-1}}{4\ 500\ \text{kg} + 1\ 000\ \text{kg}}$ $= 19.09\ \text{m.s}^{-1}$

The final velocity of the truck and car is $19.09~{
m m.s}^{-1}$ in the initial direction of the vehicles.

b.

$$egin{aligned} & EK_i = EK_{i1} + EK_{i2} \ & = rac{1}{2} \cdot {m_1} \cdot {v_{i1}}^2 + rac{1}{2} \cdot {m_2} \cdot {v_{i2}}^2 \ & = rac{1}{2} \cdot \left({m_1} \cdot {v_{i1}}^2 + {m_2} \cdot {v_{i2}}^2
ight) \ & = rac{1}{2} \Big(4 \,\, 500 \,\, \mathrm{kg} imes \left(20 \,\, \mathrm{m.s}^{-1}
ight)^2 + 1 \,\, 000 \,\, \mathrm{kg} imes \left(15 \,\, \mathrm{m.s}^{-1}
ight)^2 \Big) \ & = 1 \,\, 012 \,\, 500 \,\, \mathrm{J} \end{aligned}$$

$$egin{aligned} & EK_f = EK_{f1} + EK_{f2} = rac{1}{2} \cdot m_1 \cdot {v_{f1}}^2 + rac{1}{2} \cdot m_2 \cdot {v_{f2}}^2 \ & ext{But } v_{f1} = v_{f2} = v_f \ & \therefore EK_f = rac{1}{2} \cdot {v_f}^2 \left(m_1 + m_2
ight) \ & = rac{1}{2} imes \left(19.09 \ ext{m.s}^{-1}
ight)^2 imes \left(4 \ 500 \ ext{kg} + 1 \ 000 \ ext{kg}
ight) \ & = 1 \ 002 \ 177.28 \ ext{J} \end{aligned}$$

- c. The total kinetic energy after the collision is less than before the collision. This is because some of the kinetic energy was converted into sound and heat and was used up in the permanent deformation of both the truck and car.
- d. This was an inelastic collision because the kinetic energy of the system was not conserved.

2.

a. Let the 4 ball be object 1 and the 2 ball be object 2. $m_1 = 200 \, \, {\rm g} = 0.2 \, \, {\rm kg}$

 $v_{i1} = 2.5 \text{ m.s}^{-1}$ $m_2 = 200 \text{ g} = 0.2 \text{ kg}$ $v_{i2} = 1.5 \text{ m.s}^{-1}$ $v_{f1} = 1.5 \text{ m.s}^{-1}$

Momentum and kinetic energy are conserved. Define the initial direction of 4 ball as positive.

$$egin{aligned} m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ &\therefore m_2 \cdot v_{f2} &= m_1 \cdot v_{i1} + m_2 \cdot v_{i2} - m_1 \cdot v_{f1} \ &\therefore v_{f2} &= rac{m_1 \cdot v_{i1} + m_2 \cdot v_{i2} - m_1 \cdot v_{f1}}{m_2} \ &= rac{0.2 ext{ kg} imes 2.5 ext{ m.s}^{-1} + 0.2 ext{ kg} imes (-1.5 ext{ m.s}^{-1}) - 0.2 ext{ kg} imes (-1.5 ext{ m.s}^{-1}) \ &0.2 ext{ kg} \end{aligned}$$

 $=2.5~\mathrm{m.s}^{-1}$

The 2 ball will travel at $2.5\,\,m.s^{-1}$ in the initial direction of the 4 ball.

b.

$$egin{aligned} & EK_i = EK_{i1} + EK_{i2} \ & = rac{1}{2} \cdot m_1 \cdot {v_{i1}}^2 + rac{1}{2} \cdot m_2 \cdot {v_{i2}}^2 \ & = rac{1}{2} \cdot \left(m_1 \cdot {v_{i1}}^2 + m_2 \cdot {v_{i2}}^2
ight) \ & = rac{1}{2} \Big(0.2 \ \mathrm{kg} imes ig(2.5 \ \mathrm{m.s}^{-1} ig)^2 + 0.2 \ \mathrm{kg} imes ig(-1.5 \ \mathrm{m.s}^{-1} ig)^2 ig) \ & = 0.85 \ \mathrm{J} \end{aligned}$$

$$egin{aligned} & EK_f = EK_{f1} + EK_{f2} \ & = rac{1}{2} \cdot m_1 \cdot {v_{i1}}^2 + rac{1}{2} \cdot m_2 \cdot {v_{i2}}^2 \ & = rac{1}{2} \cdot ig(m_1 \cdot {v_{i1}}^2 + m_2 \cdot {v_{i2}}^2ig) \ & = rac{1}{2} ig(0.2 \ ext{kg} imes ig(-1.5 \ ext{m.s}^{-1}ig)^2 + 0.2 \ ext{kg} imes ig(2.5 \ ext{m.s}^{-1}ig)^2ig) \ & = 0.85 \ ext{J} \end{aligned}$$

Therefore, $EK_i = EK_f$. Kinetic energy was conserved and so the collision was elastic.

Back to Exercise 2.2

Unit 2: Assessment

1.

- a. The total momentum of an isolated system before a collision is equal to the total momentum of the system after a collision, or the total momentum of an isolated system is constant, or the total change in momentum of an isolated system is zero.
- b. Let the thief's car be object 1 and the truck be object 2. $m_1=1\,\,000\,\,{
 m kg}$

 $v_{i1} = 40 \text{ m.s}^{-1} \text{ east}$ $m_2 = 5 \ 000 \text{ kg}$ $v_{i2} = 20 \text{ m.s}^{-1} \text{ west}$ $v_{f1} = v_{f2} = v_f$ Define east as positive.

$$egin{aligned} m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ dots \cdot m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= v_f \cdot (m_1 + m_2) \ dots \cdot v_f &= rac{m_1 \cdot v_{i1} + m_2 \cdot v_{i2}}{(m_1 + m_2)} \ &= rac{1\ 000\ ext{kg} imes 40\ ext{m.s}^{-1} + 5\ 000\ ext{kg} imes (-20\ ext{m.s}^{-1}) \ 1\ 000\ ext{kg} + 5\ 000\ ext{kg}} \ &= -10\ ext{m.s}^{-1} \end{aligned}$$

The velocity of the car and truck after the collision is $10\ m.s^{-1}$ west.

$$egin{aligned} \Delta p &= m_1 \cdot v_{f1} - m_1 \cdot v_{i1} \ &= 1 \,\,000\,\,\mathrm{kg} imes (-10\,\,\mathrm{m.s^{-1}}) - 1\,\,000\,\,\mathrm{kg} imes 40\,\,\mathrm{m.s^{-1}} \ &= -50\,\,000\,\,\mathrm{kg.m.s^{-1}} \ F_{nett} &= rac{\Delta p}{\Delta t} \ &= rac{-50\,\,000\,\,\mathrm{kg.m.s^{-1}}}{0.5\,\,\mathrm{s}} \ &= -100\,\,000\,\,\mathrm{N} \end{aligned}$$

The occupants of the car would be subject to a force of 100 000 N. Therefore, the collision could result in a fatal injury for the occupants of the car. The occupants of the truck would be subject to an equal and opposite force. Therefore, the collision could result in a fatal injury for the occupants of the truck as well.

2.

a. Let the car be object 1 and the taxi be object 2. $m_1 = 1\ 000 \text{ kg}$ $v_{i1} = 20 \text{ m.s}^{-1} \text{ west}$ $m_2 = 1\ 500 \text{ kg}$ $v_{i2} = 25 \text{ m.s}^{-1} \text{ east}$ $v_{f1} = v_{f2} = v_f$ Define west as positive.

$$egin{aligned} &m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ dots \cdot m_1 \cdot v_{i1} + m_2 \cdot v_{i2} = v_f \cdot (m_1 + m_2) \ dots \cdot v_f &= rac{m_1 \cdot v_{i1} + m_2 \cdot v_{i2}}{(m_1 + m_2)} \ &= rac{1\ 000\ ext{kg} imes 20\ ext{m.s}^{-1} + 1\ 500\ ext{kg} imes (-25\ ext{m.s}^{-1}) \ 1\ 000\ ext{kg} + 1\ 500\ ext{kg} \ &= -7\ ext{m.s}^{-1} \end{aligned}$$

The velocity of the car and truck after the collision is 7 m.s^{-1} east.

b. The learner is correct in that the magnitude of the force experienced by both vehicles will be the same while their direction will be opposite. This is Newton's third law of motion which states that all forces between two objects exist in equal magnitude and opposite direction: if object A exerts a force F_A on object B, then B simultaneously exerts a force F_B on A, and the two forces are equal in magnitude and opposite in direction: $F_A = -F_B$.

$$\begin{split} \text{Impulse} &= \Delta p \\ &= m_2 v_{f2} - m_2 v_{i2} \\ &= 1\ 500\ \text{kg} \times - \left(7\ \text{m.s}^{\text{-1}} - (-25\ \text{m.s}^{\text{-1}})\right) \\ &= 27\ 000\ \text{kg.m.s}^{\text{-1}} \end{split}$$
 Or 27\ 000 kg.m.s^{-1} east

d.

$$\begin{split} EK_i &= EK_{i1} + EK_{i2} \\ &= \frac{1}{2} \cdot m_1 \cdot v_{i1}{}^2 + \frac{1}{2} \cdot m_2 \cdot v_{i2}{}^2 \\ &= \frac{1}{2} \cdot \left(m_1 \cdot v_{i1}{}^2 + m_2 \cdot v_{i2}{}^2 \right) \\ &= \frac{1}{2} \left(1\ 000\ \text{kg} \times \left(20\ \text{m.s}{}^{-1} \right)^2 + 1\ 500\ \text{kg} \times \left(-25\ \text{m.s}{}^{-1} \right)^2 \right) \\ &= 668\ 750\ \text{J} \\ EK_f &= EK_{f1} + EK_{f2} \\ &= \frac{1}{2} \cdot m_1 \cdot v_{f1}{}^2 + \frac{1}{2} \cdot m_2 \cdot v_{f2}{}^2 \\ &= \frac{1}{2} \cdot \left(m_1 \cdot v_{f1}{}^2 + m_2 \cdot v_{f2}{}^2 \right) \\ &= \frac{1}{2} \left(1\ 000\ \text{kg} \times \left(-7\ \text{m.s}{}^{-1} \right)^2 + 1\ 500\ \text{kg} \times \left(-7\ \text{m.s}{}^{-1} \right)^2 \right) \\ &= 61\ 250\ \text{J} \end{split}$$

Therefore $EK_i \neq EK_f$. Kinetic energy was not conserved and so the collision was inelastic.

3.

a. Let the 3 kg object be object 1 and the 6 kg object be object 2. Define object 1's initial direction as positive.

$$egin{aligned} m_1 &= 3 \, \mathrm{kg} \ v_{i1} &= 6 \, \mathrm{m.s^{-1}} \ v_{f1} &= -2 \, \mathrm{m.s^{-1}} \ m_2 &= 6 \, \mathrm{kg} \ v_{i2} &= 0 \, \mathrm{m.s^{-1}} \ m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ \mathrm{But} \, v_{i2} &= 0 \, \, \mathrm{m.s^{-1}} \ &\therefore m_1 \cdot v_{i1} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ &\therefore v_{f2} &= rac{m_1 \cdot v_{i1} - m_1 \cdot v_{f1}}{m_2} \ &\therefore v_{f2} &= rac{3 \, \mathrm{kg} \times 6 \, \mathrm{m.s^{-1}} - 3 \, \mathrm{kg} \times (-2 \, \mathrm{m.s^{-1}})}{6 \, \mathrm{kg}} \ &= 4 \, \mathrm{m.s^{-1}} \end{aligned}$$

The 6 $\,\rm kg$ object will move at 4 $\,\rm m.s^{-1}$ in the initial direction of the 3 $\rm kg$ object.

b.

$$egin{aligned} EK_i &= EK_{i1} + EK_{i2} \ &= rac{1}{2} \cdot {m_1} \cdot {v_{i1}}^2 + rac{1}{2} \cdot {m_2} \cdot {v_{i2}}^2 \ &= rac{1}{2} \cdot ig({m_1} \cdot {v_{i1}}^2 + {m_2} \cdot {v_{i2}}^2ig) \ &= rac{1}{2} ig(3 \, \mathrm{kg} imes ig(6 \, \mathrm{m.s}^{\text{-}1}ig)^2 + 6 \, \mathrm{kg} imes ig(0 \, \mathrm{m.s}^{\text{-}1}ig)^2ig) \ &= 54 \, \mathrm{J} \end{aligned}$$

$$egin{aligned} & EK_f = EK_{f1} + EK_{f2} \ & = rac{1}{2} \cdot m_1 \cdot v_{f1}{}^2 + rac{1}{2} \cdot m_2 \cdot v_{f2}{}^2 \ & = rac{1}{2} \cdot \left(m_1 \cdot v_{f1}{}^2 + m_2 \cdot v_{f2}{}^2
ight) \ & = rac{1}{2} igg(3 \, \mathrm{kg} imes igg(-2 \, \mathrm{m.s}{}^{-1}igg)^2 + 6 \, \mathrm{kg} imes igg(4 \, \mathrm{m.s}{}^{-1}igg)^2igg) \ & = 54 \, \mathrm{J} \end{aligned}$$

Therefore $EK_i = EK_f$. Kinetic energy was conserved and so the collision was elastic.

4.

C.

a. Let the locomotive be object 1 and the wagon be object 2. Define north as positive. $m_1 = 4 \,\,500\,\,{\rm kg}$

$$\begin{aligned} v_{i1} &= 2.5 \text{ m.s}^{-1} \text{ north} \\ m_2 &= 7 \ 000 \text{ kg} \\ v_{i2} &= 0 \text{ m.s}^{-1} \\ v_{f2} &= 1.5 \text{ m.s}^{-1} \text{ north} \end{aligned}$$
$$m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \\ \text{But} \ v_{i2} &= 0 \text{ m.s}^{-1} \\ \therefore m_1 \cdot v_{i1} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \\ \therefore v_{f1} &= \frac{m_1 \cdot v_{i1} - m_2 \cdot v_{f2}}{m_1} \\ &= \frac{4 \ 500 \text{ kg} \times 2.5 \text{ m.s}^{-1} - 7 \ 000 \text{ kg} \times (1.5 \text{ m.s}^{-1})}{4 \ 500 \text{ kg}} \\ &= 0.17 \text{ m.s}^{-1} \end{aligned}$$

The locomotive will have a velocity of $0.17\ {\rm m.s^{-1}}$ north after the collision.

b. The conservation of momentum

$$\begin{split} EK_i &= EK_{i1} + EK_{i2} \\ &= \frac{1}{2} \cdot m_1 \cdot v_{i1}{}^2 + \frac{1}{2} \cdot m_2 \cdot v_{i2}{}^2 \\ &= \frac{1}{2} \cdot \left(m_1 \cdot v_{i1}{}^2 + m_2 \cdot v_{i2}{}^2 \right) \\ &= \frac{1}{2} \left(4\ 500\ \text{kg} \times \left(2.5\ \text{m.s}{}^{\text{-}1} \right)^2 + 7\ 000\ \text{kg} \times \left(0\ \text{m.s}{}^{\text{-}1} \right)^2 \right) \\ &= 14\ 062.5\ \text{J} \\ EK_f &= EK_{f1} + EK_{f2} \\ &= \frac{1}{2} \cdot m_1 \cdot v_{f1}{}^2 + \frac{1}{2} \cdot m_2 \cdot v_{f2}{}^2 \\ &= \frac{1}{2} \cdot \left(m_1 \cdot v_{f1}{}^2 + m_2 \cdot v_{f2}{}^2 \right) \\ &= \frac{1}{2} \left(4\ 500\ \text{kg} \times \left(0.17\ \text{m.s}{}^{\text{-}1} \right)^2 + 7\ 000\ \text{kg} \times \left(1.5\ \text{m.s}{}^{\text{-}1} \right)^2 \right) \\ &= 7\ 940.03\ \text{J} \end{split}$$

Therefore $EK_i \neq EK_f$. Kinetic energy was not conserved and so the collision was inelastic. d. Now $v_{f1} = v_{f2} = v_f$.

$$egin{aligned} m_1 \cdot v_{i1} + m_2 \cdot v_{i2} &= m_1 \cdot v_{f1} + m_2 \cdot v_{f2} \ & ext{But} \; v_{i2} &= 0 \; ext{m.s}^{-1} \ &dots \; m_1 \cdot v_{i1} &= v_f \cdot (m_1 + m_2) \ &dots \; v_f &= rac{m_1 \cdot v_{i1}}{(m_1 + m_2)} \ &= rac{4\; 500 \; ext{kg} imes 2.5 \; ext{m.s}^{-1}}{4\; 500 \; ext{kg} + 7\; 000 \; ext{kg}} \ &= 0.98 \; ext{m.s}^{-1} \end{aligned}$$

The locomotive will have had a velocity of $0.98\,\,m.s^{-1}$ north if the coupling had worked.

Back to Unit 2: Assessment

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SUBJECT OUTCOME III MECHANICS: DEFINE, INTERPRET AND APPLY PRINCIPLES OF WORK, POWER AND ENERGY



Subject outcome

Subject outcome 2.3: Define, interpret and apply principles of work, power and energy



Learning outcomes

- Define work and identify examples where no mechanical work is done.
- Define and calculate work done on an object and/or system for vertical and horizontal displacement.
- Define and apply mechanical power.



Unit 1 outcomes

By the end of this unit you will be able to:

- Define work and nett work.
- Calculate the amount of work done on an object when its displacement is either vertical or horizontal.



Unit 2 outcomes

By the end of this unit you will be able to:

- Define and identify conservative forces.
- Define and identify non-conservative forces.
- State and apply the work-energy theorem.



By the end of this unit you will be able to:

• Define and apply mechanical power.

Unit 1: Energy and work

DYLAN BUSA AND LINDA PRETORIUS



What you should know

Before you start this unit, make sure you can:

- Draw and use vector diagrams. Refer to level 3 subject outcome 2.2 unit 1 if you need help with this.
- Define energy and describe the different forms that energy can take. Refer to <u>level 2 subject outcome</u>
 <u>2.3 unit 1</u> if you need help with this.
- Describe and explain the concept of conservation of energy. Refer to <u>level 2 subject outcome 2.3 unit 2</u> if you need help with this.
- Identify and determine the components of a force. Refer to <u>level 3 subject outcome 2.2 unit 2</u> if you need help with this.
- Construct and interpret free-body diagrams. Refer to <u>level 3 subject outcome 2.2 unit 3</u> if you need help with this.

Introduction

Think for a moment about the definition of 'work'.

'Work' is a word we use often in everyday conversation. We go to work. We work at getting better at something. Some things are hard work! But in physics, the word 'work' has a very specific meaning.

In the previous subject outcome on momentum, we learnt that the **momentum** of a system is always conserved (the total energy in an isolated system is always constant). In elastic collisions, however, the **kinetic energy** of the system is also conserved – it is the same before and after the collision. In inelastic collisions, some of the kinetic energy of the system gets transformed into other kinds of energy.

In both cases, however, we have objects applying a force to other objects and transferring their energy. The greater the energy of the object, the greater the force it can apply. This is the basic idea of work. The more energy an object has, the more work it can do. If object A transfers some of its energy to object B, the energy of object A decreases by the same amount of energy that object B increases by, and object A has done work on object B.

What is work?

Work, in physics, is a measure of **energy transfer** that occurs when an object is moved over a distance by an external force. In other words, when a **force** acts upon an object to cause a **displacement** of the object, it is said that **work** was done upon the object. The three important ingredients are 'force', 'displacement' and 'cause'. In order for a force to do work on an object, there must be a displacement and it must be the force that causes the displacement.

Can you think of at least three everyday occurrences where work is done? Stop and see if you can list them.

Here are some examples:

- a horse pulling a wagon
- a shopper pushing a shopping trolley
- a learner lifting a full bags of books onto their shoulder.

In each case there is a force exerted upon an object that causes that object to be displaced. In each case, **the displacement is in the same direction as the nett applied force**.





Time required: 10 minutes

What you need:

- a piece of paper
- a pen or pencil

What to do:

Think about the following situations. Is work being done? Write down whether work is done or not done in each case.

- 1. A weightlifter lifting a weight off the ground.
- 2. A weightlifter pushing a weight off his chest.
- 3. A weightlifter holding a weight stationary above his head.
- 4. Pushing against a wall.
- 5. A book falling off a table.
- 6. An aeroplane accelerating through the sky.
- 7. Pulling a box across a room.
- 8. A ball rolling down a hill.

9. A waiter carrying a tray across a room at a constant speed.

What did you find?

- 1. If a weightlifter lifts a weight off the ground, he is applying a force parallel with the direction of motion of the object and the object is moved over a certain distance, so **work is done**.
- 2. If a weightlifter pushes a weight off his chest, he is applying a force parallel with the direction of motion of the object and the object is moved over a certain distance, so **work is done**.
- 3. If a weightlifter holds a weight stationary over his head, there is no displacement and so **no work is done**. Even though he is exerting a force to counter-balance the force of gravity, because there is no displacement of the weight, no work is done.
- 4. If you push against a wall, you certainly exert a force against the wall and the wall exerts a force against you. However, neither the wall nor you move over a distance and so **no work is done**.
- 5. If a book falls off a table, the force of gravity acts in the direction of the motion of the book and the book is moved over a certain distance, so **work is done**. Even through the velocity of the book is not constant during the fall, the force being applied is constant.
- 6. If an aeroplane accelerates through the sky, the force parallel to the motion of the aircraft is applied to it and it moves over a certain distance, so **work is done**.
- 7. If a box is pulled across a room, there is a force parallel to the direction of motion of the box and the box moves over a distance and so **work is done**.
- 8. When a ball rolls down a hill, there is a component of the force of gravity acting in parallel to the motion of the ball and the ball moves over a distance, therefore, **work is done**.
- 9. There is a force on the tray, but this is upwards to counteract the force of gravity. There is a displacement of the tray as it moves across the room. However, the force does not cause the displacement. There is no component of the lifting force in the direction of the displacement and so the force is not causing the displacement. Therefore, **no work is done**.

Scenario 9 in activity 1.1 was a tricky one which we will explore in more detail later. For now, it is important to remember that for work to be done, the force or at least a component of the force, must be in the direction of the displacement.

The work done can be expressed mathematically as $W = F\Delta x \cos \theta$ where:

- W is the work done (measured in newton metres or joules 1 N.m = 1 J)
- *F* is the force acting on the object (measured in newtons)
- $\cdot \Delta x$ is the displacement (measured in metres) note that Δx can also be represented as d
- θ is the angle between the force and the displacement vector (hence $\cos \theta$ gives us the component of the force that acts in parallel with the displacement (see figure 1).



Figure 1: Calculating the magnitude of the horizontal component

If the force is parallel to the direction of motion, and in the same direction, then $\theta = 0^{\circ}$, $\cos \theta = 1$ and $W = F\Delta x \cos \theta$ becomes simply $W = F\Delta x$ (see figures 2 and 3).



Figure 2: Cases of work being done where the force is parallel and in the same direction as the displacement



$W = F\Delta \vec{x} \cos 0^{\circ} = F\Delta \vec{x}(1) = F\Delta \vec{x}$

Figure 3: The force and direction of motion are perfectly parallel

If the force is parallel to the direction of motion, but in the opposite direction, then $\theta = 180^{\circ}$, $\cos \theta = -1$ and $W = F\Delta x \cos \theta$ become $W = F\Delta x(-1)$ (see figure 4). We get a negative value for work. It is important to note that work is not a vector and so only has a **magnitude**. If negative work is done, energy is lost (rather than gained) and the object exerting the force gains this energy. For example, if you try to push a car uphill by applying a force up the slope and instead the car rolls down the hill, you are doing negative work on the car. Another way of saying this is that the car is doing positive work on you.



$W = F\Delta \vec{x} \cos 180^{\circ} = F\Delta \vec{x}(-1) = -F\Delta \vec{x}$

Figure 4: Force and direction of motion are parallel but in opposite directions

If the force and the direction of displacement are perpendicular, then $\theta = 90^{\circ}$, $\cos \theta = 0$ and $W = F\Delta x \cos \theta$ become $W = F\Delta x(0) = 0$. Therefore, no work is done.

Before we move on, let's examine scenario 9 of the waiter carrying the tray from activity 1.1 in more detail. If the waiter and tray started at rest, there would be a brief period of time when the waiter would have had to apply a force to the tray to get it to start moving (i.e. to accelerate it). However, once the tray is at its constant speed, it will stay in its straight-line motion at this constant speed without a forward force. If the only force exerted upon the tray during this constant speed stage of its motion is upward, then no work is done on the tray during this period. Remember, a vertical force does not do work on a horizontally displaced object.



If the force is perfectly parallel but in the opposite direction to the direction of motion, then $\theta = 180^{\circ}$ and so $\cos \theta = \cos 180^{\circ} = -1$ and $W = F\Delta x \times (-1) = -F\Delta x$.

If the force and the direction of displacement are perpendicular then $\theta = 90^{\circ}$ and so $\cos \theta = 0$ and $W = F\Delta x \cos \theta = F\Delta x \times 0 = 0$.

We will only deal with occasions where $heta=0^\circ$ or $heta=180^\circ.$



Example 1.1

A bicycle is travelling along a straight horizontal road. The cyclist applies a force of 50 N through the pedals and chain in the direction in which she is travelling, and is speeding up. While she is accelerating, she covers a distance of 65 m. Calculate the work done by the cyclist.

Solution

F = 50 N $\Delta x = 65 \text{ m}$

$$heta x = 05 \,\,\mathrm{m}$$
 $heta = 0^{\circ}$

 0° the applied force of 50 N is in the direction of travel

We need to calculate the work done.

 $egin{aligned} W &= F\Delta x\cos heta\ &= 50\,\,\mathrm{N} imes 65\,\,\mathrm{m} imes \cos0^\circ\ &= 50\,\,\mathrm{N} imes 65\,\,\mathrm{m} imes 1\ &= 3\,\,250\,\,\mathrm{N.m} \end{aligned}$

3 250 N.m or 3 250 J of work is done. This means that the energy of the bicycle (and rider) increased.



Example 1.2

The same cyclist as above applies her brakes to slow down. The force applied by the brakes opposite to her direction of travel is 45 N and she travels 50 m while braking. Calculate the work done by the brakes.

Solution

F = 45 N

 $\Delta x = 50 \,\,\mathrm{m}$

 $heta = 180^\circ$ the applied force of 45 N is in the opposite direction to the direction of travel

We need to calculate the work done.

 $W = F\Delta x\cos heta$

 $=45\,\,\mathrm{N} imes50\,\,\mathrm{m} imes\cos180^\circ$

$$=45\,\,\mathrm{N} imes50\,\,\mathrm{m} imes(-1)$$

$$= -2\ 250\ \mathrm{N.m}$$

 $-2~250~{
m N.m}$ or $-2~250~{
m J}$ of work is done. This means that that energy of the bicycle (and rider) decreased.

Nett work

Hardly ever is there only a single force acting on an object. In activity 1.2, we will investigate some simple cases of this.



Time required: 15 minutes

What you need:

- \cdot a pen or pencil
- \cdot a piece of paper

What to do:

Study the following free-body diagrams and determine which forces are doing work and how much work these forces are doing. Then determine the overall work being done.

1. A 15 N force is applied to push a block across a frictionless surface for a displacement of 10.0 m to the right.



2. A 15 N frictional force slows a block moving to the right to a stop after a displacement of 10.0 m.







What did you find?

- 1. The gravitational and normal forces act perpendicular to the displacement. Therefore, they do no work. The applied force acts parallel to the displacement; therefore, it does work. The work done by this force is $W = F\Delta x \cos \theta = 15 \text{ N} \times 10 \text{ m} \times \cos 0^{\circ} = 150 \text{ J}$. This is the only force doing work, so the nett work is 150 J.
- 2. The gravitational and normal forces act perpendicular to the displacement. Therefore, they do no work. The frictional force acts parallel (but in the opposite direction) to the displacement; therefore, it does work. The work done by this force is $W = F\Delta x \cos \theta = 15 \text{ N} \times 10 \text{ m} \times \cos 180^{\circ} = -150 \text{ J}.$ This is the only force doing work, so the nett work is -150 J.
- 3. The gravitational and normal forces act perpendicular to the displacement. Therefore, they do no work. The applied force acts parallel to the displacement; therefore, it does work. The work done by this force is $W = F\Delta x \cos \theta = 15 \text{ N} \times 10 \text{ m} \times \cos 0^{\circ} = 150 \text{ J}$. The frictional force acts parallel (but in the opposite direction) to the displacement; therefore, it does work too. The work done by this

force is $W = F\Delta x \cos \theta = 15 \text{ N} \times 10 \text{ m} \times \cos 180^{\circ} = -150 \text{ J}$. Therefore, the nett work is 150 J + (-150 J) = 0 J. In other words, no nett work is done. This is a very important result!

- 4. The gravitational and normal forces act perpendicular to the displacement. Therefore, they do no work. There is no force (or component of a force) acting in parallel to the displacement. Therefore, no work is done.
- 5. The gravitational and normal forces act parallel to the displacement. Therefore, they do work. The work done by the upward tension force is $W = F\Delta x \cos \theta = 29.4 \text{ N} \times 10 \text{ m} \times \cos 0^{\circ} = 294 \text{ J}$. The gravitational force acts parallel (but in the opposite direction) to the displacement; therefore, it does work. The work done by this force is

 $W = F\Delta x \cos \theta = 15 \text{ N} \times 10 \text{ m} \times \cos 180^{\circ} = 3 \text{ kg} \times 9.8 \text{ m.s}^{-2} \times 10 \text{ m} \times (-1) = -294 \text{ J}$. Therefore, the nett work is 294 J + (-294 J) = 0 J. In other words, no nett work is done.

Example 1.3

How much work is done by an applied force to lift a 15 kg block 12.0 m vertically at a constant speed?

Solution

 (\bigcirc)

Firstly, if the block is lifted at a constant velocity, this means that there is no acceleration. Therefore, the force lifting the block will be equal but opposite to the force of gravity. The applied force must, therefore, be equal to the weight of the block.

```
m = 15 \text{ kg}

g = 9.8 \text{ m.s}^{-2}

h = 12.0 \text{ m}

F = mg = 15 \text{ kg} \times 9.8 \text{ m.s}^{-2} = 147 \text{ N}

W = F\Delta x \cos \theta

But \Delta x = h and \theta = 0^{\circ}

\therefore W = F \times h

= 147 \text{ N} \times 12 \text{ m}

= 1 764 \text{ J}
```



Example 1.4

A force of 500 N is accelerating a car down a flat straight road. Friction is working against the motion of the car. A frictional force of 350 N acts to oppose the motion. Calculate the nett work done on the car to move it forward 50 m.

Solution

The applied force is 500 N and the frictional force is 350 N acting in the opposite direction to the car's motion. There is no work done by gravity or the normal force because these are perpendicular to the car's motion. The road is flat.

 $F_{applied} = 500 \ \mathrm{N}$ $F_{fric} = 350 \,\,\mathrm{N}$ $\Delta x = 50 \,\,\mathrm{m}$ $W_{applied} = F\Delta x\cos heta$ $= 500\,\,\mathrm{N} \times 50\,\,\mathrm{m} \times \cos0^\circ$ $= 25 \ 000 \ \mathrm{J}$ $W_{friction} = F\Delta x\cos heta$ $= 350~N\times 50~m\times cos\, 180^\circ$ $= -17 \,\, 500 \,\, {\rm J}$ $W_{net} = W_{applied} + W_{friction}$ $= 25 \ 000 \ J - 17 \ 500 \ J$ $= 7 \ 500 \ J$ Alternatively, we can calculate the total work by first calculating the nett force on the car. Define forward as positive $F_{applied} = 500 \,\,\mathrm{N}$ $F_{friction} = -350$ N

Both forces act in the same horizontal plane so the vector addition is simple and does not require calculating any horizontal components.

 $F_{nett} = F_{applied} + F_{friction} = 500 \ \mathrm{N}{-}\,350 \ \mathrm{N} = 150 \ \mathrm{N}$

Neither gravitational force nor the normal force have horizontal components – they act perpendicular to the motion.




Exercise 1.1

- 1. A traveller carries a 250 N suitcase up four flights of stairs (a total height of 10.5 m) and then pushes it with a horizontal force of 50 N at a constant speed of $0.25 \text{ m}.\text{s}^{-1}$ for a horizontal distance of 50 m on a frictionless surface. How much work does the traveller do on the suitcase during this entire trip?
- 2. How much work is done by the force required to raise a 3 000 N lift five floors vertically at a constant speed? The vertical distance between floors is 5 m high.
- 3. A learner with a mass of 60 kg runs up three flights of stairs in 15 s, covering a vertical distance of 10 m. Determine the amount of work done by the learner to elevate her body to this height if her speed was constant.

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Work is done on an object when a force acts on an object to cause a displacement of the object.
- Work is expressed mathematically as $W = F\Delta x \cos \theta$ where θ is the angle between the force and the displacement vector.
- Nett work is the work done by the nett force acting on an object that causes a displacement of the object.

Unit 1: Assessment

Suggested time to complete: 30 minutes

1. Before beginning its initial descent, a roller coaster car is pulled up the first hill to a high initial height. Work is done on the car to achieve this initial height. A roller coaster designer is considering three different incline angles at which to drag the 2 500 kg car to the top of a 70 m high hill. In each case, the force applied to the car will be applied parallel to the hill. Which of the following options would require the most work?

Option	Angle	Force	Distance
А	35°	$1.12 imes 10^4~ m N$	105 m
В	45°	$1.39 imes 10^4$ N	84.60 m
С	55°	$1.61 imes 10^4~ m N$	73.04 m

- 2. A 120 kg rugby player does push-ups by applying a force to elevate his centre-of-mass by 30 cm. Calculate the number of push-ups that he needs to do to do at least $10 \ 000 \text{ J}$ of work. Assume no air resistance.
- 3. Car A and car B both have a mass of 950 kg. Car A can accelerate from rest to 100 km/h in 16 s. Car B can accelerate from rest to 100 km/h in 3.5 s.
 - a. To accelerate to 100 km/h, which engine will do more work. Justify your answer with calculations. Assume that both cars accelerate in a straight line. Ignore the effects of air resistance and friction.
 - b. What is this work?

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

- 1. Work done up the steps:
 - $W = F\Delta x \cos heta$
 - $=250\,\,\mathrm{N} imes10.5\,\,\mathrm{m} imes\cos0^\circ$
 - $= 2 \,\, 625 \,\, \mathrm{J}$

Work done pushing: $W = F\Delta x \cos \theta$ $= 50 \text{ N} \times 50 \text{ m} \times \cos 0^{\circ}$ = 2 500 JTotal work = 2 625 J + 2 500 J = 5 125 J2. $W = F\Delta x \cos \theta$ $= 3 0000 \text{ N} \times (5 \times 5) \text{ m}$ = 75 000 J3. As speed is constant, the learner is only doing work enough to equal the force of gravity. $W = F\Delta x \cos \theta$

= 60 kg × 9.8 m.s⁻² × 10 m × cos 0° = 5 880 J

Back to Exercise 1.1

Unit 1: Assessment

1. The angle in the table is the incline angle. The angle θ in the work equation is the angle between F and Δx . If the force is parallel to the incline and the direction of the displacement is parallel to the incline, then the angle θ in the work equation is 0° . For this reason, $W = F\Delta x \cos 0^{\circ} = F\Delta x$.

$$\begin{split} W &= F\Delta x\cos\theta\\ \text{A: } 35^\circ\text{:}\ W &= 1.12\times10^4\ \text{N}\times105\ \text{m} = 1\ 176\ 000\ \text{J}\\ \text{B: } 45^\circ\text{:}\ W &= 1.39\times10^4\ \text{N}\times84.60\ \text{m} = 1\ 175\ 940\ \text{J}\\ \text{C: } 55^\circ\text{:}\ W &= 1.61\times10^4\ \text{N}\times73.04\ \text{m} = 1\ 175\ 944\ \text{J} \end{split}$$

While there are small differences in the work done, given the amount of work these can largely be put down to rounding errors. It really does not make much difference which option is used. The total work required to get the car to the top will be about the same (about $1 \ 176 \ 000 \ J$)

2. Each push up requires the player to do work only against gravity to lift his body.

 $W=F\Delta x\cos heta$

= 120 kg \times 9.8 m.s⁻² \times 0.3 m = 352.8 J 000 J

 $\frac{10\ 000\ J}{352.8\ J}=28.34$

Therefore, the player will need to do at least 29 push-ups to do $10\,\,000\,\,\mathrm{J}$ of work.

3.

a. In both cases, the force accelerating the cars is parallel and in the same direction to the displacement. Therefore, $\theta = 0^{\circ}$ and $\cos \theta = 1$.

$$egin{aligned} W &= F\Delta x \ \mathrm{But} \; F &= ma \ dots \; W &= ma\Delta x \ \mathrm{But} \; a &= rac{v_f - v_i}{t} \ dots \; W &= rac{m(vf - vi)}{t}\Delta x \ \mathrm{But} \; \Delta x &= s = rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f)t \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f) \ dots \; W &= rac{m(v_f - v_i)}{t} imes rac{1}{2}(v_i + v_f) \ dots \; W &= rac{m(v_f - v_i)}{t} imes V \ u_f \ u_f$$

Therefore, as both cars have the same mass, the same initial velocity and the same final velocity, both engines will need to do the same amount of work.

$$egin{aligned} &m=950 \ \mathrm{kg} \ &v_i=0 \ \mathrm{m.s}^{-1} \ &v_f=100 \ \mathrm{km/h}=27.78 \ \mathrm{m.s}^{-1} \ &W=rac{1}{2}m(v_f{-}v_i)(v_i+v_f) \ &=rac{1}{2}m(v_f)(v_f) \ &=rac{1}{2}m(v_f)^2 \ &=rac{1}{2} imes 950 \ \mathrm{kg} imes (27.78 \ \mathrm{m.s}^{-1})^2 \ &=366 \ 570.99 \ \mathrm{J} \end{aligned}$$

Notice how the total work done by the engines is equal to the change in kinetic energy. We will investigate this idea in more detail in unit 2.

Refer back to Refer to level 2 subject outcome 2.1 if you need to on how to determine acceleration.

Back to Unit 1: Assessment

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Unit 2: Energy-work theorem

DYLAN BUSA AND LINDA PRETORIUS



What you should know

Before you start this unit, make sure you can

- Define and calculate work. Refer to <u>unit 1</u> if you need help with this.
- Define mechanical energy.
- Define kinetic energy.
- Define gravitational potential energy.
- \cdot Calculate kinetic energy ($E_k=rac{1}{2}mv^2$).
- Calculate gravitational potential energy ($E_p = mgh$).
- Refer to level 2 subject outcome 2.3 if you need help with any of these other concepts.

Introduction

In level 2 subject outcome 2.3, we learnt that an object has potential energy due to its position above the earth or its position in earth's field of gravity. This is **gravitational potential energy**. If you let go of the object, it will fall, and it can do work on your toe! We also learnt that the amount of this gravitational potential energy is proportional to the mass of the object and its height. The heavier the object is or the greater the height the object falls from, the more work it can do on your toe.

We calculate gravitational potential energy as $E_p = mgh$.

Note that another form of potential energy is **elastic potential energy**. This is the energy stored in a compressed spring or a stretched elastic.

We also learnt that an object has energy due to its motion, more specifically, its velocity. This is **kinetic energy**. The faster an object is moving, the more energy it has and the more work it can do on other objects. We also learnt that the amount of kinetic energy is proportional to the object's mass and the square of its velocity.

We calculate kinetic energy as $E_k=rac{1}{2}mv^2.$

The mechanical energy of an object is simply the sum of its potential energy and kinetic energy and is, therefore, a measure of the total amount of work that the object can do.

We calculate mechanical energy as $E_{mechanical} = E_p + E_k$.

Work and energy

We know from level 2 subject outcome 2.3 that mechanical energy is conserved in a closed system. In other words, if you drop a ball from a height, all the ball's gravitational potential energy at this height is converted to kinetic energy when it hits the ground. In other words, in figure 1, the ball's mechanical energy at point A is equal to the ball's mechanical energy at point B. At point A, the ball's mechanical energy is all gravitational potential energy. At point B it has all been converted into kinetic energy.



Figure 1: Conservation of mechanical energy

But is this always the case? Is mechanical energy always conserved? Think of a car travelling on a flat road. What happens to its mechanical energy if the engine stops?

Our experience tells us that the car's mechanical energy is not conserved. It will eventually roll to a stop. In this case, its gravitational potential energy is not changing (it is on a flat road) but its kinetic energy certainly is. We know why the car eventually stops and loses all its mechanical energy. Forces like friction and the rolling resistance of the tyres on the ground always exert a force in the opposite direction to the car's motion.

Mechanical energy is only conserved when all the forces acting on an object are **conservative forces**. Examples of conservative forces include gravity and spring forces. They are called conservative forces precisely because they result in the conservation of mechanical energy as illustrated in figure 1.

If **non-conservative forces** act on the object, then its total mechanical energy will change. Non-conservative forces, as the name suggests, do not result in the conservation of mechanical energy. Examples include friction, air resistance, rolling resistance, any applied force and tension forces.

If a body is acted on by non-conservative forces, its mechanical energy will either increase or decrease. If

positive work is done, then the object will gain energy. If negative work is done, then the object will lose energy. This gain or loss in energy can be in the form of potential energy, kinetic energy, or both.



Take note!

Conservative forces result in the **conservation of mechanical energy**. The object's energy simply changes form from kinetic to potential or vice versa. Conservation forces are sometimes called internal forces. The work done by a conservative force only depends on the starting and ending points of the motion not the path taken.

Non-conservative forces result in the **change of mechanical energy**. The object gains or loses mechanical energy. Non-conservative forces are sometimes called external forces. The work done by non-conservative forces depends on the path taken.



Activity 2.1: Mechanical energy is conserved

Time required: 10 minutes

What you need:

- a pen or pencil
- a piece of paper

What to do:

Study the following simple scenarios, and in each case:

- identify what kinds of forces are involved in doing the work
- · identify whether mechanical energy is being conserved or not
- if mechanical energy is conserved, state whether it is being from kinetic energy to potential energy or vice versa
- if mechanical energy is not conserved, state whether there is a change in the kinetic energy, potential energy or both.
- 1. A ball falls from a height of 3 m in the absence of air resistance.
- 2. A stationary ball just after it is kicked along the ground.
- 3. A skier glides from one point to another down an undulating friction-free icy slope.
- 4. A cricket ball is traveling vertically upwards.
- 5. A horizontally moving hammer just after it strikes a nail into a block of wood.
- 6. The end of a compressed spring as it is released.

What did you find?

- 1. The only force doing work on the ball is gravity. Therefore, mechanical energy is conserved. As the ball loses height it gains velocity. Potential energy is converted into kinetic energy.
- 2. The applied force of the foot causes the ball to accelerate and hence gain kinetic energy. As there is no change in the potential energy of the ball, the ball gains mechanical energy. Thus, a non-

conservative applied force changes the kinetic energy of the ball.

- 3. The only force doing work on the skier is gravity. Therefore, mechanical energy is conserved. As the skier loses height on the downhills, she gains velocity. As she loses velocity on the uphills, she gains height. Potential energy is converted into kinetic energy and vice versa but as her initial height is greater than her final height, the overall conversion is from potential energy to kinetic energy.
- 4. The only force doing work on the ball is gravity. Therefore, mechanical energy is conserved. As the ball gains height, it loses velocity. Kinetic energy is converted into potential energy.
- 5. The hammer is moving horizontally so there is no change in potential energy. The equal and opposite force of the nail acting on the hammer causes the hammer to stop or gain negative velocity. Therefore, there is a change in the hammer's kinetic energy and its mechanical energy is not conserved.
- 6. The only force doing work on the end of the spring is the compression force of the spring. The spring changes from a compressed state to a relaxed state changing the velocity of the end of the spring. Thus, the conservative potential spring energy is transformed into kinetic energy.



Take note!

That mechanical energy is not conserved when non-conservative forces act on an object does not mean that **total energy** is not conserved. Total energy is always conserved. Non-conservative forces simply mean that **mechanical energy** is not conserved in a particular system which implies that some energy is transferred in a process that isn't reversible, such as losses to heat.

The work-energy theorem

Consider the simple case of sliding a box across the floor (see figure 2). There is an applied force pulling the box to the right and a frictional force acting in the opposite direction. We will define to the right as positive.



Figure 2: Pulling a box across the floor

- If $F_{applied} > F_{friction}$, there is a **nett force to the right**. The box will accelerate to the right. It will **gain velocity**. **Nett positive work** will be done on the box.
- If $F_{applied} < F_{friction}$, there is a **nett force to the left**. The box will accelerate to the left (its acceleration will be negative). It will **lose velocity**. **Nett negative work** will be done on the box.
- If $F_{applied} = F_{friction}$, there is no nett force. The box will experience no acceleration. There will be no change in its velocity. If it is stationary, it will remain stationary. If it is moving it will remain moving in this direction at a constant velocity. No **nett work** will be done on the box. This does not mean that the applied force and the frictional force are not individually doing work, just that the nett work on the box

will be zero.

Now consider the simple case of a box being lifted into the air (see figure 3). There is an applied force lifting the box up and a gravitational force acting in the opposite direction. We



- If $F_{applied} > F_{gravity}$, there is a **nett force up**. The box will accelerate up. It will **gain velocity**. **Nett positive work** will be done on the box.
- If $F_{applied} < F_{gravity}$, there is a **nett force down**. The box will accelerate down (its acceleration will be negative). It will **lose velocity**. **Nett negative work** will be done on the box.
- If $F_{applied} = F_{gravity}$, there is **no nett force**. The box will experience no acceleration. There will be no change in its velocity. If it is stationary, it will remain stationary. If it is moving it will remain moving in this direction at a constant velocity. No **nett work** will be done on the box. This does not mean that the applied force and the gravitational force are not individually doing work, just that the nett work on the box will be zero.

This leads us to the conclusion that when a **nett force** does work on an object, then there is always a **change in the kinetic energy of the object**. This is because the object experiences an acceleration and therefore a change in velocity.

This leads to the **work-energy theorem**.

The work-energy theorem states that the work done on an object by the nett force is equal to the change in its kinetic energy.

$$W_{nett} = \Delta E_k = E_{kf} - E_{ki}$$

To understand what this means in practice, it is helpful to look at some examples.

Example 2.1

Question adapted from Everything Science Grade 12 Worked example 6 page 232
A 2 kg brick is dropped from a height of 15 m. Calculate the work that has been done on the brick
between the moment it is released and the moment when it hits the ground. Assume that air
resistance can be neglected.
Solution
The brick falls freely with only conservative forces acting on it, so energy is conserved. We know that
the work done is equal to the brick's change in kinetic energy. The brick has no kinetic energy at the
moment it is dropped, because it is stationary. When the brick hits the ground, all the brick's potential
energy is converted to kinetic energy.

$$m = 2 \text{ kg}$$

 $h = 15 \text{ m}$
 $E_P = mgh$
 $= 2 \text{ kg} × 9.8 \text{ m.s}^{-2} \times 15 \text{ m}$
 $= 294 \text{ J}$
But the brick had 294 J of potential energy when it was released and 0 J of kinetic energy.
Therefore $E_{ki} = 0 \text{ J}$ and $E_{kf} = 294 \text{ J}$.
 $W_{nett} = \Delta E_k = E_{kf} - E_{ki}$
 $= 294 \text{ J}$
Example 2.2
A 980 kg car traveling on a flat road skids to a stop from an initial velocity of 27 m.s^{-1} . The car's brakes

Solution

There is no change in the potential energy of the car. Therefore, we can ignore any change in potential energy. We will define the direction of the car's initial motion as positive.

apply a 7 $\,650\,$ N of frictional force. What is the stopping distance of the car?

$$\begin{split} m &= 980 \text{ kg} \\ v_i &= 27 \text{ m/s}^{-1} \\ v_f &= 0 \text{ m/s}^{-1} \\ F_{nett} &= 7 \text{ 650 N} \\ E_{Ki} &= \frac{1}{2} \text{mv}^2 \\ &= \frac{1}{2} 980 \text{ kg} \times (27 \text{ m.s}^{-1})^2 \\ &= 357 \text{ 210 J} \\ E_{Kf} &= 0 \text{ J} \\ W_{nett} &= \Delta E_k = E_{kf} - E_{ki} \\ \therefore W_{nett} &= 0 \text{ J} - 357210 \text{ J} \\ &= -357 \text{ 210 J} \\ mode that nett work is negative i.e. the car loses mechanical energy} \\ W_{nett} &= F_{nett} \Delta x \cos \theta \\ \therefore \Delta x &= \frac{W_{nett}}{F_{nett}} \times \cos \theta \\ &= \frac{-357 \text{ 210 J}}{7 \text{ 650 N}} \times \cos 180^\circ \\ &= 46.69 \text{ m} \\ \end{split}$$
The car will stop in 46.69 m.

We can use the work-energy theorem to generate an expression to calculate the total work done by all the non-conservative forces in a system.

We know that $F_{nett} = F_{conservative} + F_{non-conservative}$ (the nett work done is the sum of the work done by all of the individual forces).

Therefore, $W_{nett} = W_{conservative} + W_{non-conservative}$.

But the work-energy theorem tells us that $W_{nett} = \Delta E_K$. Therefore, $\Delta E_K = W_{conservative} + W_{non-conservative}$.

But the work done by conservative forces can be expressed as ΔPE . Remember, if you lift an object at a constant velocity, the conservative force (gravitational force) acts in the opposite direction to the motion (negative work) but its magnitude is equal to the gain in potential energy of the object.

Therefore, $\Delta E_K = -\Delta E_P + W_{non-conservative}$.

Rearranging this equation, we finally get that $W_{non-conservative} = \Delta E_K + \Delta E_P$.

As we can see from the equation above, when the non-conservative forces oppose the motion of the object, the work done by these forces is negative, causing a decrease in the mechanical energy of the system. When the non-conservative forces do positive work, energy is added to the system. If the sum of the non-conservative forces is zero then $\Delta E_K = -\Delta E_P$ and mechanical energy is conserved.



Take note!

For situations where there is a change in the kinetic and/or the potential energy of an object, the total work done by the non-conservative forces to enact this change can be calculated using: $W_{non-conservative} = \Delta E_K + \Delta E_P.$ Once the total work of these forces is known, the total nett non-conservative forces or the distance over which the work is done can be calculated.



Example 2.3

Question adapted from Everything Science Grade 12 Worked example 9 page 242

Consider the situation where a football player slides to a stop on level ground. Using energy considerations, calculate the distance a 65.0 kg football player slides, given that his initial speed is 6.00 m.s^{-1} and the force of friction against him is a constant 450 N.

Solution

We know that the frictional force will act in the opposite direction to the player's initial motion to reduce his kinetic energy to zero. Most of this kinetic energy will be converted to heat, deformation of the grass and some sound. While the total energy of the system is conserved, the mechanical energy of the player is not. We know this because there is a frictional non-conservative force acting on the player.

Because the frictional force acts to reduce the player's mechanical energy, we know that it does negative work on the player; $\theta = 180^{\circ}$. Therefore, $\cos \theta = -1$.

Because the player is on level ground, there is no change in the potential energy. Hence $\Delta E_P = 0$.

We know that $W_{non-conservative} = \Delta E_K + \Delta E_P$. But $\Delta E_P = 0$. Therefore, $W_{non-conservative} = \Delta E_K$. This is basically the original work-energy theorem.

$$\begin{split} m &= 65 \text{ kg} \\ v_i &= 6 \text{ m.s}^{-1} \\ F_{friction} &= 450 \text{ N} \\ W_{non-conservative} &= \Delta E_K \\ \therefore F_{friction} \Delta x \cos \theta &= E_{Kf} - E_{Ki} \\ \therefore -F_{friction} \Delta x &= 0 - E_{Ki} \\ \therefore \Delta x &= \frac{-E_{Ki}}{F_{friction}} \\ \therefore \Delta x &= \frac{-E_{Ki}}{F_{friction}} \\ &= -\frac{1}{2} \times 65 \text{ kg} \times \frac{(6 \text{ m.s}^{-1})^2}{450 \text{ N}} \\ &= 2.60 \text{ m} \end{split}$$

The player will stop after 2.60 m.

Exercise 2.1

1. Consider the falling and rolling motion of a 2 kg ball in the following two resistance-free situations.

In one situation, the ball falls off the top of the platform to the floor. In the other situation, the ball rolls from the top of the platform along the staircase-like pathway to the floor.



For each situation:

- a. indicate what types of forces are doing work upon the balls
- b. indicate whether the energy of each ball is conserved and explain why
- c. calculate values for all unknowns a k in the diagram.
- 2. At the end of a roller coaster ride, the 6 000 kg train of cars (including passengers) is slowed from a speed of $_{20 m.s}^{-1}$ to a speed of $_{5 m.s}^{-1}$ over a distance of $_{20 m}$ up an incline that has a total vertical height of $_{2 m}$. Calculate the braking force required to slow the train of cars by this amount.
- 3. A girl on roller blades has an initial velocity of $10 \text{ m}.\text{s}^{-1}$. She has a mass of 50 kg and coasts straight up a hill that is 2.5 m height.



- a. What distance does she travel up the hill?
- b. What is her speed at the top of the hill if the total frictional force acting on her is 1.5 N?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- · Conservative forces (or internal forces) conserve mechanical energy.
- Non-conservative forces (or external forces) do not conserve mechanical energy.
- The work done by conservative forces does not depend on the path taken.

- The work done by non-conservative forces depends on the path taken.
- The work-energy theorem states that total nett work is equal to the change in kinetic energy; therefore $W_{nett} = \Delta E_K = E_{Kf} E_{Ki}$.
- The work-energy theorem can be rearranged to calculate the total non-conservative work done on an object as $W_{non-conservative} = \Delta E_K + \Delta E_P$.

Unit 2: Assessment

Suggested time to complete: 35 minutes

- A shopping trolley full of groceries is sitting at the top of a 2.0 m hill. The trolley rolls down the hill and hits a stump at the bottom of the hill. Upon impact, a 250 g can of baked beans flies horizontally out of the shopping cart and hits a parked car with an average force of 500 N. How deep a dent is made in the car (i.e. over what distance does the 500 N force act upon the can of beans before bringing it to a stop)?
- 2. A rope is attached to a 40.0 kg crate to pull it up a frictionless incline at constant speed to a height of 3 m. A diagram of the situation and a free-body diagram are shown below.



Note that the force of gravity has two components (parallel and perpendicular components); the parallel component balances the applied force, and the perpendicular component balances the normal force.

- a. Of the forces acting on the crate, which ones do work?
- b. Based upon the types of forces acting upon the system and their classification as conservative and non-conservative forces, is mechanical energy conserved? Explain.
- c. Calculate the amount of work done upon the crate.

The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Exercise 2.1

1.

- a. The only force doing work on the ball is gravity.
- b. Since gravity is a conservative force, the total mechanical energy of the ball is conserved. Therefore, the total mechanical energy of the ball remains 100 J.
- c. a 50 J. If the ball has half its original potential energy, the other half must be been converted into kinetic energy.

$$b - 7.07 \, \mathrm{m.s^{-1}} \, E_K = rac{1}{2} m v^2 \therefore v = \sqrt{rac{100}{2}} = \sqrt{50}.$$

 $c-50\,$ J. The ball is at the same height, therefore, its potential energy is still $50\,$ J.

d – 50 J. There is no resistance, therefore, its kinetic energy is still 50 J.

 $e-7.07~m.s^{-1}.$ There is no resistance, therefore, its velocity is still $7.07~m.s^{-1}$ 50 J.

 $f-0\,\;J.$ The ball is on the ground, therefore, it has no potential energy.

g – 100 J. The ball is on the ground, therefore, all its potential energy has been converted to kinetic energy.

h - 10 m.s⁻¹
$$E_K = rac{1}{2}mv^2$$
 \therefore $v = \sqrt{rac{200}{2}} = \sqrt{100}.$

 ${\rm i-0}\,$ J. The ball is on the ground, therefore, it has no potential energy.

j – 100 J. The ball is on the ground, therefore, all its potential energy has been converted to kinetic energy.

$$extsf{k-10 m.s}^{-1} E_K = rac{1}{2} m v^2 \therefore v = \sqrt{rac{200}{2}} = \sqrt{100}.$$

2.

 $m = 6 \,\, 000 \,\, {
m kg}$ $vi = 20 \,\, {
m m.s^{-1}}$ $vf = 5 \,\, {
m m.s^{-1}}$ $\Delta x = 20 \,\, {
m m}$ $h = 2 \,\, {
m m}$

 $W_{non-conservative} = \Delta E_K + \Delta E_P$ Where:

$$\cdot \ \ W_{non-conservative} = F_{brake} \Delta x \cos heta = -F_{brake} \Delta x$$

$$\cdot \quad \Delta E_K = E_{Kf} - E_{Ki} = rac{1}{2} m v_f^2 - rac{1}{2} m v_i^2 = rac{1}{2} m (v_f^2 - v_i^2)$$

$$\cdot \Delta E_P = mgh$$

$$egin{aligned} & \therefore -F_{brake}\Delta x = rac{1}{2}m(v_f^2 - v_i^2) + mgh \ & \therefore F_{brake} = -rac{\left(rac{1}{2}m(v_f^2 - v_i^2) + mgh
ight)}{\Delta x} \ & = -rac{\left(rac{1}{2} imes 6\ 000\ ext{kg}\left((5\ ext{m.s}^{-1})^2 - (20\ ext{m.s}^{-1})^2
ight) + 6\ 000\ ext{kg} imes 9.8\ ext{m.s}^{-2} imes 2\ ext{m})} \ & = 50\ 370\ ext{N} \end{aligned}$$

Therefore, the braking force required is 50 370 N.

3.

a. We can use trigonometry to calculate the distance up the hill.

$$\sin 25^{\circ} = \frac{2.5 \text{ m}}{d}$$
$$\therefore d = \frac{2.5 \text{ m}}{\sin 25^{\circ}}$$
$$= 5.92 \text{ m}$$

b. The value calculated in a. is Δx in calculation of work.

$$m = 60 \,\, {
m kg} \ \Delta x = 5.92 \,\, {
m m} \ F_{friction} = 1.5 \,\, {
m N} \ h = 2.5 \,\, {
m m} \ v_i = 10 \,\, {
m m.s}^{-1}$$

We need to calculate the final velocity of the girl at the top of the hill. In this case, she will lose kinetic energy and gain potential energy.

 $W_{non-conservative} = \Delta E_K + \Delta E_P$ Where:

 $\cdot \hspace{0.1 cm} W_{non-conservative} = F_{friction} \Delta x \cos heta = -F_{friction} \Delta x$

•
$$\Delta E_K = E_{Kf} - E_{Ki} = \frac{1}{2}mv_f^2 - \frac{1}{2}mv_i^2 = \frac{1}{2}m(v_f^2 - v_i^2)$$

 $\cdot \ \Delta E_P = mgh$

$$\begin{array}{l} \therefore -F_{friction} \Delta x = \frac{1}{2}m(v_{f}^{2} - v_{i}^{2}) + mgh \\ \therefore -F_{friction} \Delta x - mgh = \frac{1}{2}m(v_{f}^{2} - v_{i}^{2}) \\ \therefore -2\frac{(F_{friction} \Delta x + mgh)}{m} = v_{f}^{2} - v_{i}^{2} \\ \therefore v_{f}^{2} = -2\frac{(F_{friction} \Delta x + mgh)}{m} + v_{i}^{2} \\ = -2\frac{(-1.5 \text{ N} \times 5.92 \text{ m}) + (50 \text{ kg} \times 9.8 \text{ m.s}^{-2} \times 2.5 \text{ m})}{50 \text{ kg}} + (10 \text{ m.s}^{-1})^{2} \\ = 51.3552 \text{ m.s}^{-2} \\ \therefore v_{f} = 7.17 \text{ m.s}^{-1} \end{array}$$

Her speed at the top of the hill will be $7.17\ m.s^{-1}.$

Back to Exercise 2.1

Unit 2: Assessment

1. The only force acting on the trolley and, therefore the tin, is gravity. Therefore, mechanical energy is conserved and all the potential energy at the top of the hill is converted to kinetic energy at the bottom of the hill.

$$m=250\,\,{
m g}=0.25\,\,{
m kg}$$

 $h=2\,\,{
m m}$
 E_P at the top of the hill $=mgh$
 $=0.25\,\,{
m kg} imes9.8\,\,{
m m.s}^{-2} imes2\,\,{
m m}$
 $=4.9\,\,{
m J}$
 $\therefore E_K$ at the bottom of the hill $=4.9\,\,{
m J}$

When the tin hits the car, it loses kinetic energy due to the force applied to it by the body of the car. This is a non-conservative force. Therefore, mechanical energy is not conserved. All the can's kinetic energy is transformed into the energy needed to dent the car. The force the car body applies to the can is 500 N in the opposite direction to the motion of the can.

 $V_f = 0 \text{ m.s}^{-1}$ F = 500 N $W_{nett} = \Delta E_K$ $\therefore F \Delta x \cos \theta = E_{Kf} - E_{Ki}$ $\therefore 500 \text{ N} \Delta x (-1) = 0 \text{ J} - 4.9 \text{ J}$ $\therefore \Delta x = \frac{4.9 \text{ J}}{500 \text{ N}}$ = 0.0098 m

Therefore, It will take $9.8~\mathrm{mm}$ for the car to stop the can. In other words, the dent will be $9.8~\mathrm{mm}$ deep.

2.

- a. Both gravity (because of the component of gravity acting parallel to the direction of motion) and the applied tension force in the rope do work. The normal force is perpendicular to the direction of motion so does no work.
- b. The applied force is non-conservative. Since it does work on the box, the mechanical energy of the box is not conserved.
- c. $W_{non-conservative} = \Delta E_K + \Delta E_P$ Where
 - $\cdot \quad \Delta E_K = 0 \, \, \mathrm{J}$ (there is no change in velocity)

$$\cdot \quad \Delta E_P = mgh$$

 $\therefore W_{non-conservative} = mgh$ = 40 kg × 9.8 m.s⁻² × 3 m = 1 176 J

The work done by the non-conservative force in pulling the box up the incline is $1\ 176\ J.$

Note that it does not matter what the angle of the incline or distance of the incline is. Because there is no change in kinetic energy, all the non-conservative work is done 'against gravity'. And the work done by gravity, being a conservative force, does not depend on the path taken; only the start and end points i.e. the change in height.

Back to Unit 2: Assessment

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Unit 3: Mechanical power

DYLAN BUSA AND LINDA PRETORIUS



What you should know

Before you start this unit, make sure you can:

- Define and calculate work and nett work. Refer to <u>unit 1</u> if you need help with this.
- State and apply the work-energy theorem. Refer to <u>unit 2</u> if you need help with this.

Introduction

In the previous units we have defined and used the concepts of energy and work. We know that work can be defined as the energy transferred by a force. But what about how quickly this energy is transferred? In other words, what about how quickly work is done?

Think of the following scenario. Two cars line up at the start of a race. One is an ordinary road car and the other is a sports car. Both cars can accelerate to 100 km/h from rest. The ordinary road car can do it in 16 s. The sports car can do it in 3.5 s. If we assume that both cars have the same mass (950 kg), which car's engine will do more work? Which car's engine is more powerful?

What is power?

You may recognise the scenario above from the first question in the unit l assessment and therefore already know that the work done by both engines (ignoring wind resistance and friction) is the same. Remember that this is based on the work-energy theorem. We know that $W_{net} = \Delta E_k$. As both cars experience exactly the same change in kinetic energy (both cars accelerate from 0 m.s^{-1} to 27.78 m.s⁻¹), the net work done is the same.

But the sports car is able to do this work far quicker. We would say that it has a more powerful engine. This everyday use of the word 'power' is very similar to the physics use. We call the rate at which work can be

done **power**. Therefore,
$$P = \frac{W}{t}$$
.

In unit 1, we calculated that the net work done by both engines is 366 570.99 J. However, what is the power of each engine?

The road car: $P = \frac{W}{t} = \frac{366\ 570.99\ \text{J}}{16\ \text{s}} = 22\ 910.69\ \text{W}$ The sports car:

$$P = \frac{W}{t} = \frac{366\ 570.99\ \text{J}}{3.5\ \text{s}} = 104\ 734.57\ \text{W}$$

The unit for power is joules per second $(\frac{J}{s} \text{ or } J_{.s}^{-1})$. However, the more common unit of power is the watt (W); $1J.s^{-1} = 1$ W.

Clearly, the sports car engine is far more powerful than the road car engine. It can do the same amount of work in a far shorter time.

Using the definition of work, we can easily derive an alternative expression for power:

 $P = \frac{W}{t} = \frac{F\Delta x\cos\theta}{t}$

If the force and the change in displacement are in the same direction then:

$$P = rac{F\Delta x}{t}$$

But $\frac{\Delta x}{t}$ is the definition of velocity. Therefore: P=Fv

Note that if the force and velocity are in the opposite direction then power will be negative. Like work, power is not a vector. It only has magnitude.





Example 3.2

Question adapted from Everything Science Grade 12 Worked example 12 page 246

A forklift lifts a crate of mass 100 kg at a constant velocity to a height of 8 m over a time of 4 s. The forklift then holds the crate in place for 20 s.

- 1. Calculate how much power the forklift exerts in lifting the crate?
- 2. How much power does the forklift exert in holding the crate in place?

Solutions

1.

$$m_{crate} = 100 \, \, \mathrm{kg}$$

$$h_{crate} = 8 \text{ m}$$

t=4 s

We need to calculate the power required to lift the crate.

$$P = \frac{W}{t}$$

$$= \frac{F\Delta x}{t}$$

$$= \frac{m_{crate} \cdot g \times h_{crate}}{t}$$

$$= \frac{100 \text{ kg} \times 9.8 \text{ m.s}^{-2} \times 8 \text{ m}}{4 \text{ s}}$$

$$= 1 960 \text{ W}$$

2. While the crate is being held, there is no displacement and so the work done is zero. If no work is done, then no power is exerted.

Example 3.3

A 55 kg athlete trains by doing stair sprints – running up a flight of stairs. From rest, she runs up three flights of stairs each 2.5 m in height in 7.2 s. If her final velocity is 7 m.s⁻¹ what is her power output?

Solution

Her total work will be the increase in her kinetic energy plus her increase in gravitational potential energy. Therefore $W = \Delta E_k + \Delta E_p$. We are told that she starts from rest and we can assume that her initial gravitational potential energy is also zero (relative to her gravitational potential energy at the top of the stairs).

$$egin{aligned} W &= \Delta E_k + \Delta E_p \ dots W &= (E_{kf} - E_{ki}) + (E_{pf} - E_{pi}) \ &= (E_{kf} - 0) + (E_{pf} - 0) \ &= rac{1}{2}mv^2 + mgh \ v &= 7 \ {
m m.s}^{-1} \ m &= 55 \ {
m kg} \ g &= 9.8 \ {
m m.s}^{-2} \ h &= 3 imes 2.5 \ {
m m} = 7.5 \ {
m m} \ W &= rac{1}{2} imes 55 \ {
m kg} (7 \ {
m m.s}^{-1})^2 + 55 \ {
m kg} imes 9.8 \ {
m m.s}^{-2} imes 7.5 \ {
m m} \ &= 5 \ 390 \ {
m J} \end{aligned}$$

She does 5 390 J of work in 7.2 s. Therefore, her power output is $P = \frac{W}{t} = \frac{5 390 \text{ J}}{7.2 \text{ s}} = 748.61 \text{ W}.$

Note that only 1 347.5 J of work went into increasing her velocity from rest to 7 $m.s^{-1}$. The bulk of the work (4 042.5 J) was required to lift her 7.5 m. This is why stair sprints are such effective workouts.

Exercise 3.1

- 1. Does the work required to lift a book to a high shelf depend on how fast you lift it? Does the power required to lift the book depend on how fast you lift it? Explain.
- 2. A winch designed to be mounted on a truck is advertised as being able to exert a 6.8×10^3 N force and to develop a power of 0.30 kW. How long would it take the truck and the winch to pull an object 15 m?
- 3. An elevator motor lifts a total mass of $1.1x10^3$ kg a distance of 40.0 m in 12.5 s at a constant velocity. How much power does the elevator motor generate during this period?
- 4. A windmill is used on a farm to pump water out of a well. The water flows up 37 m above the well water level, before being directed to the dam. The water is pumped at 2 m.s^{-1} .



- a. Calculate how much energy is necessary to pump 90 kg of water out of the ell to a height of 37 $\,\rm m$
- b. It is necessary to pump at least $90 \ \mathrm{kg}$ of water per minute. What is the minimum power that the windmill must produce?
- c. The farmer wants to modernise the farm. The farmer decides to buy a $0.5 \ kW$ petrol water pump. Will the petrol water pump be able to pump at the required rate?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Power is the rate at which work is done.
- The unit for power is joules per second (J.s⁻¹), however, the more common unit of power is the watt (W); $1J.s^{-1} = 1$ W.
- If force and velocity are in the opposite direction, then power will be negative.
- Like work, power is not a vector. It only has magnitude.

Unit 3: Assessment

Suggested time to complete: 45 minutes

Question 1 is adapted from NC(V) Grade 12 Paper 1 February 2019 question 8

1. A farmer irrigates his crops using water drawn from a storage tank. The water falls by gravity down through the pipe at a rate of $0.05 \text{ m}^3.\text{s}^{-1}$. The water falls through a height of 8 m. The mass of 1 m^3 of water is $1 \times 10^3 \text{ kg}$. Ignore the effects of friction.



- a. Calculate the mass of water that flows through the downpipe each second.
- b. State the principle of conservation of mechanical energy.
- c. Using this principle, calculate the kinetic energy of the mass of water in a. above as the water reaches the bottom of the downpipe.
- d. The farmer decides to install a small hydro-electric generator at the bottom of the downpipe. Assume that the generator is able to convert 75% of the maximum kinetic energy gained by the water into electricity. Calculate the electrical power output of the generator.
- 2. A cable car has a mass of 3 000 kg when fully laden. An electrical motor is used to pull the cable car up from the lower base station (at 366 m above sea level) to the top of the mountain, which is 1 066 m

above sea level. The motor can deliver a maximum of $88\,$ kW. It takes seven minutes for the cable car to move from the lower base station to the top of the mountain.

- a. Define power.
- b. If the motor operates at maximum power output, calculate the electrical energy used by the motor to pull the cable car.
- c. Calculate the gain in gravitational potential energy of the fully laden car when it moves from the bottom station to the top station. Compare your answer with the value calculated in the previous question and explain the difference.
- d. Should the braking mechanism of the fully laden car fail while it is at rest at the top of the station, at what speed will the car strike the bottom station, if during the descent, the car loses 6×10^6 J of mechanical energy in doing work against friction?
- 3. Hikers climb to the top of a waterfall that is 948 m high. 6×10^4 kg of water falls from the top of the waterfall each minute. (Ignore air resistance and any other frictional forces).
 - a. Define gravitational potential energy.
 - b. Calculate the gravitational potential energy of $6 \times 10^4 \ \text{kg}$ of water at the top of the waterfall, relative to the bottom of the waterfall.
 - c. The gravitational potential energy stored in this water could be used to turn a turbine connected to an electrical generator at the bottom of the falls. Describe the energy conversion that takes place in a generator.
 - d. If 80% of the power that the water has owing to its gravitational potential energy is converted into electrical power, calculate the output power of the generator?
 - e. What would be the average kinetic energy of a kilogram of water when it reaches the bottom of the waterfall?

The <u>full solutions</u> are at the end of the unit.

Unit 3: Solutions

Exercise 3.1

- 1. Work is purely a function of force and distance. It does not matter how fast you lift the book. The total amount of work done is the same. Power, however, is the rate at which work is done. Therefore, the faster your raise the book, the more power you deliver.
- 2.

$$P = 0.3 \text{ kW} = 300 \text{ W}$$

$$F = 6.8 × 10^3 \text{ N}$$

$$W = F \Delta x$$

= 6.8 × 10³ N × 15 m
= 102 000 J

$$P = \frac{W}{t}$$

∴ $t = \frac{W}{P}$
= $\frac{102 \ 000 \text{ J}}{300 \text{ W}}$
= 340 s

3. During this period, there is no change in the velocity of the lift. Therefore, the lift does not gain any kinetic energy. The only work done is to overcome the weight of the lift. This work is exactly equal to

the weight of the lift. There is acceleration hence no net force.

$$m = 1.1 \times 10^3 \text{ kg}$$

 $\Delta x = 40 \text{ m}$
 $t = 12.5 \text{ s}$
 $W = F\Delta x$
 $= mg\Delta x$
 $= 1.1 \times 10^3 \text{ kg} \times 9.8 \text{ m.s}^{-1} \times 40 \text{ m}$
 $= 431\ 200 \text{ J}$
 $P = \frac{W}{t}$
 $= \frac{431\ 200 \text{ J}}{12.5 \text{ s}}$
 $= 34\ 496 \text{ W}$

4.

a. The water is given potential energy and kinetic energy as its velocity is increased from $0 m.s^{-1} to 2 m.s^{-1}$.

$$h = 37 \text{ m}$$

$$v = 2 \text{ m.s}^{-1}$$

$$m = 90 \text{ kg}$$

$$W = \Delta E_k + \Delta E_p$$

$$= \frac{1}{2}mv^2 + mgh$$

$$= \frac{1}{2} \times 90 \text{ kg} \times (2 \text{ m.s}^{-1})^2 + 90 \text{ kg} \times 9.8 \text{ m.s}^{-2} \times 37 \text{ m}$$

$$= 32 814 \text{ J}$$

$$W = 32 814 \text{ J}$$

$$W = 32 814 \text{ J}$$

$$t = 60 \text{ s}$$

$$P = \frac{W}{t}$$

$$= \frac{32 814 \text{ J}}{60 \text{ s}}$$

$$= 546.9 \text{ W}$$

c. The 0.5 kW pump will only be able to produce 500 W of power. Therefore, it will be insufficient to pump 90 kg of water a minute as this requires a power output of 546.9 W.

Back to Exercise 3.1

b.

Unit 3: Assessment

1.

- a. $0.05~m^3$ flows every second. If $1~m^3$ has a mass of 1~000~kg, then $0.05~m^3$ will have a mass of $1~000~kg.m^{-3}\times0.05~m^3=50~kg.$
- b. The total mechanical energy (sum of gravitational potential energy and kinetic energy) in an isolated system remains constant as long as the system is free of friction and other non-conservative forces.

c.

$$a = 9.8 ext{ m.s}^{-1}$$

 $s = 8 ext{ m}$
 $v_i = 0 ext{ m.s}^{-1}$
 $v_f^2 = v_i^2 + 2as$
 $= 2 imes 9.8 ext{ m.s}^{-2} imes 8 ext{ m}$
 $= 156.8 ext{ m}^2. ext{ s}^{-2}$
 $\therefore v_f = 12.52 ext{ m.s}^{-1}$
 $v = 12.52 ext{ m.s}^{-1}$
 $m = 50 ext{ kg}$
 $E_k = rac{1}{2}mv^2$
 $= rac{1}{2} imes 50 ext{ kg} imes (12.52 ext{ m.s}^{-1})^2$

 $= 3 \ 918.76 \ J$

Because the water falls freely (without friction), we know that all the potential energy at the top of the pipe is converted into kinetic energy at the bottom of the pipe. Therefore, the total work done is 3 918.76 J.

This work is done every second. Therefore, the total power output is 3 918.76 W. However, only 75% of this is converted to electrical power. Therefore, the electrical power output is 3 918.76 W \times 0.75 = 2 939.07 W.

2.

d.

a. Power is the rate at which work is done.

b.

$$P = 88 \text{ kW} = 88 000 \text{ W}$$

$$t = 7 \text{ minutes} = 420 \text{ s}$$

$$P = \frac{W}{t}$$
but work is equal to energy

$$\therefore E = P \times t$$

$$= 88000 \text{ W} \times 420$$

$$= 36.96 \times 10^6 \text{ J}$$
c.

$$m = 3 000 \text{ kg}$$

$$h = 1 066 \text{ m} - 366 \text{ m} = 700 \text{ m}$$

$$\Delta E_p = mgh$$

 $= 3 \,\,000 \,\, {
m kg} imes 9.8 \,\, {
m m.s}^{-2} imes 700 \,\, {
m m} \ = 20.58 imes 10^6 \,\, {
m J}$

There is a difference of $36.96 \times 10^6 \text{ J} - 20.58 \times 10^6 \text{ J} = 16.38 \times 10^6 \text{ J}$. The difference is due to energy 'lost' in overcoming friction and air resistance as well as the energy required to initially accelerate the car to its constant velocity.

d. The total energy available is the car's gravitational potential energy of 20.58×10^6 J. If 6×10^6 J is lost doing work against friction then the total energy available to be converted into kinetic energy is 14.58×10^6 J.

$$E = rac{1}{2}mv^2 \ dots v^2 = rac{2E}{m} \ = rac{2 imes 14.58 imes 10^6 \ {
m J}}{3\ 000 \ {
m kg}} \ dots v = 98.59 \ {
m m.s}^{-1}$$

This is equivalent to $355~\mathrm{km/h}$

3.

- a. Gravitational potential energy is the energy a body possesses owing to its position above a particular reference point (particularly the earth's surface).
- b.

d.

 $egin{aligned} m &= 6 imes 10^4 \ \mathrm{kg} \ h &= 948 \ \mathrm{m} \ E_p &= mgh \ &= 6 imes 10^6 \ \mathrm{kg} imes 9.8 \ \mathrm{m.s^{-2}} imes 948 \ \mathrm{m} \ &= 5.574 x 10^8 \ \mathrm{J} \end{aligned}$

c. As the water falls, its gravitational potential energy is converted into kinetic energy. Inside the turbine, the kinetic energy of the water is converted to mechanical energy of the turbine which is then converted to electrical energy.

 $6\times10^4~kg$ of water falls from the top of the waterfall each minute. Therefore, the falling water can do $5.574\times10^8~J$ every minute or 60 s. However, only 80% of this is converted.

$$egin{aligned} P &= rac{W}{t} imes 80\% \ &= rac{5.574 imes 10^8 \,\, \mathrm{J}}{60 \,\, \mathrm{s}} imes 0.8 \ &= 7.432 imes 10^6 \,\, \mathrm{W} \ E_p &= E_k \end{aligned}$$

 $m = 1 ext{ kg}$ $h = 948 ext{ m}$ $E_p = 1 ext{ kg} imes 9.8 ext{ m.s}^{-2} imes 948 ext{ m}$ $= 9 ext{ 290 J}$

Therefore, the average kinetic energy would be 9 290 J.

Back to Unit 3: Assessment

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SUBJECT OUTCOME IV MECHANICS: STATE, EVALUATE AND APPLY MECHANICAL ADVANTAGE

Subject outcome

Subject outcome 2.4: State, evaluate and apply mechanical advantage



 (\forall)

Learning outcomes

- Define torque and identify its application in simple machines.
- Describe and identify the mechanical advantage in a wheel and axle.
- Define a pulley and distinguish between fixed and movable pulleys.
- Identify and describe the mechanical advantage of a pulley system.
- Define gears and gear trains and distinguish between the use of gears and pulleys.
- Analyse and apply law (equation) of simple machines on pulleys (hoist) to determine the mechanical advantage.



Unit 1 outcomes

By the end of this unit you will be able to:

- Define torque.
- Identify the application of torque in simple machines.
- Describe and identify the mechanical advantage in a wheel and axle.



Unit 2 outcomes

By the end of this unit you will be able to:

- Define a pulley.
- Distinguish between fixed and movable pulleys.
- Identify and describe the mechanical advantage of a pulley system.
- Analyse and apply law (equation) of simple machines on pulleys (hoist) to determine the mechanical advantage.



By the end of this unit you will be able to:

- Define gears and gear trains.
- Distinguish between the use of gears and pulleys.

Unit 1: Wheels and axles and their mechanical advantage

LEIGH KLEYNHANS



What you should know

Before you start this unit, make sure you can:

• Identify, describe, and apply principles of simple machines and mechanical advantage. See <u>level 2</u> subject outcome 2.4 to revise this.

Introduction

You have learnt that simple machines are devices that can be used to multiply an applied force – often at the expense of the distance through which the force is applied. Levers, gears, pulleys, wedges, and screws are some examples of machines. Energy is still conserved for these devices because a machine cannot do more work than the energy put into it. However, machines can reduce the input force that is needed to perform the job. You have also learnt that the ratio of output to input force for any simple machine is called its mechanical advantage (MA).

What is torque?

The ability of a force to rotate a lever around a fixed point is measured by a quantity called torque.

Torque depends on the size of the force and the lever arm. The perpendicular distance from the axis of rotation to a line drawn along the direction of the force is called the lever arm (the distance, d, from the fixed axis of rotation to the point at which the force is applied).



Figure 1: A diagram showing a force (F) being applied to a lever with a fixed axis of rotation, with d representing the lever arm

Application of torque in simple machines

To illustrate how torque is used in simple machines, look at figure 2 of a pet flap door.



Figure 2: A pet flap door

The force F is applied by the dog perpendicular to the pet-flap door. The lever arm is the distance, *d*, from the dog's nose to the hinge. If the dog pressed on the door at a higher point, the lever arm would be shorter. As a result, the dog would need to exert a greater force to apply the same torque.

How easily an object rotates depends not only on how much force is applied but also on where the force is applied.

The further the force is from the axis of rotation, the easier it is to rotate the object and the more torque is produced.

Torque is directly proportional to the force applied (effort) and the length of the lever arm. This can be represented mathematically using the following formula:

au = r ot FWhere: au is the torque r ot is the length of the lever arm F is the force or effort



- 1. Tape the latch of a hinged door so that the door will open when you push without turning the door handle.
- 2. Using only a force perpendicular to the door, push the door open several times by applying a force at different distances from the hinge.
- 3. Compare the relative effort required to open the door when pushing near the edge to that required when pushing near the hinged side of the door.

What did you find?

You should have found that the longer the lever arm (the distance from the hinge to where you applied the force), the less effort (force) was required to open the door. When the lever arm is short, a greater effort (force) is required to rotate the door.

Note

Watch this video to find out more about torque. (Note that you are not required to do calculations for assessment.)

Torque (Duration: 07.02)





The <u>full solutions</u> are at the end of the unit.

Wheels and axles

The wheel and axle is a machine consisting of a wheel attached to a smaller axle so that these two parts rotate together because a force is transferred from one to the other.



Figure 3: A wheel and axle system

The wheel and axle can be viewed as a version of the lever, with a drive force applied tangentially to the perimeter of the wheel and a load force applied to the axle, balanced around the hinge which is the fulcrum.

The simple machine called a wheel and axle refers to any assembly formed by two disks, or cylinders, of different diameters mounted so they rotate together around the same axis. The thin rod which needs to be turned is called the axle and the wider object fixed to the axle, is called the wheel. A tangential force applied to the outer edge of the large disk (greater circumference = longer distance) can exert a larger force on a load attached to the outer edge of the axle (smaller circumference = shorter distance), achieving mechanical advantage. A heavy load can be lifted with a small force.

Circumference is proportional to radius, thus by varying the radii of the axle and/or wheel, mechanical advantage may be gained. The larger the ratio, the greater the multiplication of force (torque) created. If friction is ignored, the ideal mechanical advantage can be expressed using the following formula:

$$IMA = rac{FB_{out}}{FA_{in}} or rac{R_{wheel \ radius}}{r_{axle \ radius}}$$



Figure 4: The ratio of radius of wheel to radius of axle determines mechanical advantage

On a powered vehicle, the engine exerts a force on the axle which has a smaller radius than the wheel. The mechanical advantage is therefore much less than 1. The force on the wheels will be less than the force on the axle. However, the friction between wheel and road is quite low, so even a small force exerted on the axle is sufficient to create a force to rotate the wheels of larger radius. The actual advantage of this system lies in the large rotational speed at which the axle is rotating creating a lesser, but effective rotational speed in the wheels.



Figure 5: The wheel and axle in a car

Types of wheel and axle machines

Depending on the point where the force is applied, wheel and axle machines can be classified into two broad categories:

1. Machines where force is applied to the wheel

In these types of machines, the effort or force is applied to the wheel causing it to rotate. The motion of the simple machine is initiated by the wheel that develops a greater amount of force on the axle. Some examples of such machines include windmills, drills, and screwdrivers.

2. Machines where force is applied to the axle

Here, the effort or the force is applied to the axle (usually at a rapid rate), and this causes the wheel to rotate at a slower rate. This means that the motion initiated by the axle gets transferred to the wheel.

Some of the real-life applications that use this type of wheel and axle simple machine include bicycles, Ferris wheels and vehicles.

Note

Watch this video to find out more about a wheel and axle system.

The Wheel and Axle (Duration: 03.24)



Exercise 1.2

- 1. In the simple machine called a wheel and axle, what is the difference between the component referred to as the wheel and the component referred to as the axle?
- 2. When a force is applied to the wheel, how do the following compare to each other?
 - a. the force on the wheel and the force on the axle
 - b. the energy of the wheel and the energy of the axle
 - c. the rotational distance of the wheel and the rotational distance of the axle
 - d. the rotational speed of the wheel and the rotational speed of the axle
- 3. How would the answers to the above questions change if the force was applied to the axle?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Torque is the measurement of the ability of a force to rotate a lever around a fixed point.
- The torque is determined by the force and the length of the lever arm.
- The shorter the lever arm, the greater the force required for rotation. The longer the lever arm, the less force required for rotation.
- Torque is applied in simple machines such as hinged doors and windows, and when using a wrench to tighten or loosen a bolt.
- A wheel and axle is a simple machine consisting of two disks or cylinders of different radii connected together; the one with the larger radius is called the wheel and the one with the smaller radius is called the axle.
- Less force is required to rotate the larger wheel (over a larger distance), therefore more force is created at the smaller axle (over less distance).
• The mechanical advantage of a wheel and axle where the force is applied to the wheel can be expressed as:

 $IMA = rac{FB_{out}}{FA_{in}} or rac{R_{wheel \ radius}}{r_{axle \ radius}}$

• In wheel and axle systems where the force is applied to the axle, there is no mechanical advantage (increase in force) but rotation of the axle rapidly over a short distance can cause the wheel to rotate at a slower speed but over a larger distance.

Unit 1: Assessment

Suggested time to complete: 10 minutes

- 1. Define torque.
- 2. Explain why it is easier to open a window by applying a force near the opening edge rather than near the hinged edge.
- 3. Explain why it is easier to fasten a screw with a screwdriver rather than with your fingers. Support your reasoning by applying the principle of mechanical advantage.
- 4. Which answer correctly describes a wheel and axle movement?
 - A. the wheel and axle move the same distance and turn with the same amount of force
 - B. the wheel moves a lesser distance than the axle but turns with greater force
 - C. the wheel moves a greater distance than the axle but turns with less force
 - D. the wheel moves, while the axle stays completely still
- 5. Explain how a windmill makes use of the principle of the wheel and axle system. Support your answer by applying the principle of mechanical advantage.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.

- a. torque
- b. the size of the force and the length of the lever arm (distance from the axis of rotation to where the force is applied)
- 2. It would be more difficult to open the door, more force will be required (the lever arm will be shorter, so to create the same amount of torque the force needs to be greater).
- 3. Increase the force with which they push the door Apply the force as far as possible from the hinge.

Back to Exercise 1.1

Exercise 1.2

- 1. The wheel has a larger radius/circumference and is bigger than the axle.
- 2.
- a. the force on the axle will be greater than that on the wheel
- b. the energy of the wheel and axle are the same
- c. the rotational distance covered by the wheel is greater than the rotational distance covered by the axle
- d. the rotational speed of the axle is greater than the rotational speed of the wheel
- 3. They would not change.

Back to Exercise 1.2

Unit 1: Assessment

- 1. Torque is the measurement of the ability of a force to rotate a lever around a fixed point.
- 2. Torque is proportional to the force applied (effort) and the length of the lever arm (the distance from the hinge to where the force is applied). This is expressed in the formula: $\tau = r \perp F$. Therefore, to create the amount of torque to open the window, the force will decrease the further it is applied from the hinge of the window.
- 3. This is explained using the mechanical advantage of a wheel and axle system. The handle of the screwdriver has a larger radius than the radius of the screw head. The mechanical advantage of a wheel and axle system is expressed as: $IMA = \frac{FB_{out}}{FA_{in}} or \frac{R_{wheel \ radius}}{r_{axle \ radius}}$. Therefore, the output force on the screw head will be greater than the input force on the handle of the screwdriver. It will be easier to turn the screwhead with a screwdriver compared to directly applying a force to the screw head.
- 4. C
- 5. The blades of a windmill have a large radius. When a small force is applied by the wind to the blades, the force on the axle will be greater, as its radius is smaller. This creates a mechanical advantage

according to the formula: $IMA = \frac{FB_{out}}{FA_{in}} or \frac{R_{wheel \ radius}}{r_{axle \ radius}}$. The increased force on the axle can be used to

create kinetic energy to pump water or make electricity.

Back to Unit 1: Assessment

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Unit 2: Pulleys and mechanical advantage

LEIGH KLEYNHANS



What you should know

Before you start this unit, make sure you can:

• Identify, describe, and apply principles of simple machines and mechanical advantage. See <u>level 2</u> subject outcome 2.4 to revise this.

Introduction¹

In this unit, you will learn how pulleys can act as simple machines and provide a mechanical advantage. Thousands of years ago and still to this day, engineers have used pulleys to make everyday tasks and large construction projects easier. They use pulleys to raise heavy objects, taking advantage of gravity to help lift an object. Engineers incorporate pulleys into a wide range of modern-day applications, such as cranes for lifting very heavy objects in construction projects, elevators, flagpoles, motors, bicycles, and boats. To gain a mechanical advantage using multiple pulleys in conjunction with motors and electronics, engineers create devices that operate with very little power in relation to how much work they perform.

What is a pulley?

A pulley is a simple machine that consists of a wheel with a groove, running along its circumference between the rims of the wheel. A rope or cable passes through the groove, which is attached to the load on one end, and the effort is applied to perform work at the other end. A fixed pulley is one that is attached to a non-moving point (an anchor) and the rope is attached to the object that needs to be raised or lowered.

1. Parts of the text in this unit were sourced from <u>Teach Engineering</u> released under a CC-BY licence.

In a movable pulley the rope is attached to a non-moving point and the pulley is attached to the object and which is able to slide along the rope.

Fixed pulleys

A fixed pulley is useful for raising an object to a level above your head. Using this type of pulley enables you to take advantage of gravity. A real-life example of this would be raising a flag on a flagpole.



Figure 1: A fixed pulley

The pulley in figure 1 redirects the force. This means that by looping a rope around the pulley and attaching the rope to an object, one pulls down on the rope to raise the object, instead of having to lift the object. Although this is a helpful and convenient use for pulleys, it has a major limitation: the force you must apply to lift the object is the same as if you were just lifting the object without the pulley. This means that a fixed pulley does not give any mechanical advantage.

By attaching a counterweight to the end of the rope that you pull, you can lessen the amount of force you must apply. Once a force is applied to either side, the system continues moving in that direction. This kind of pulley system is used in lifts. The lift has a cable attached to it that goes up, around a pulley, then comes down and attaches to a counterweight. The motor that moves the lift uses much less power since the counterweight provides a gravitational force.

Movable pulleys

In a movable pulley system, the rope is attached to a fixed (non-moving) point, the pulley is attached to the object that you want to move, and the other end of the rope is left free (see figure 2). By pulling on the rope, the pulley moves, and the object is raised.



Figure 2: A movable pulley

This type of system is good if you are trying to raise an object located below you to your level. In a variation, if both sides of a movable pulley system are fixed and the rope is taut between the fixed points, the system becomes like a wheel and axle because the object can ride along the rope if a force is applied to it (for example, a zip line).



The mechanical advantage of using multiple pulleys

How can pulleys make our lives easier? Pulleys are powerful simple machines. They can change the direction of a force, which can make it much easier for us to move something. If we want to lift an object that weighs 10 kg to a height of 1 m, we can lift it straight up or we can use a pulley where we can pull down on one end to lift the object up. It is much easier to use the pulley because, as long as we weigh more than 10 kg, we can just hang onto the end of the rope and take advantage of gravity, so our weight provides all the necessary force to lift the object.

Pulleys can also provide us with a mechanical advantage when we use several together and more rope. This process lessens the amount of force required to lift something.



Activity 2.1: Demonstrate how using more rope can increase mechanical advantage

Time required: 10 minutes

What you need:

- · 3 people
- \cdot a 6 m piece of rope
- · 2 brooms

What to do:

- Tie the rope to one of the brooms (broom 1) and wrap the rope around the other broom (broom 2). Have two people stand about a metre apart each holding a broom, and try to keep the brooms separated while the third person pulls on the free end of the rope; it should be a difficult task to pull the broom sticks together.
- 2. Next, wrap the rope around each of the brooms again. Try again to pull the brooms together.
- 3. Repeat the process again (see figure 3).



What did you find?

The more times you wrap the rope around the brooms, the easier it is for the third person to pull the brooms together. This is an example showing the power of mechanical advantage.

Combining multiple pulleys decreases the amount of force necessary to move an object by increasing the amount of rope used to raise the object. The mechanical advantage (MA) of a pulley system is equal to the number of ropes supporting the movable load. (Do not count ropes that are only used for redirecting the force.)

MA = number of load supporting ropes

A single pulley therefore has a mechanical advantage of 1, but using multiple pulleys will proportionally increase the mechanical advantage.



Figure 4: A pulley system with a mechanical advantage of 2



Figure 5: A pulley system with a mechanical advantage of 4 because it has four load-supporting rope segments

A system of multiple pulleys that creates a significant mechanical advantage is referred to as a block and tackle. Engineers combine many pulleys into a pulley system that significantly reduces the amount of force required to lift an object. They often use pulley systems to move extremely heavy objects. A block and tackle system may take a lot of cable or rope, but a human using enough pulleys could lift several tons. Engineers use the block and tackle along with motors and electronics to create modern devices that operate with very low power requirements, such as cranes and elevators.







The <u>full solutions</u> are at the end of the unit.

Activity 2.2: Investigate the mechanical advantage of using multiple pulleys

Time required: 10 minutes

What you need:

- a plastic milk bottle filled with sand
- rope or string

What to do:

- 1. Tie the rope to a strong bar (or rod) above your head (a playground swing works well).
- 2. Run the rope through the handle of the milk bottle and then over the bar overhead.
- 3. Raise the milk bottle by pulling the rope down. Note how difficult it is to do this.
- 4. Next, run the rope through the milk bottle and over the bar again. Raise the milk bottle again.
- 5. Continue to loop the rope through the milk bottle and over the bar noting how much easier it becomes to raise the milk bottle . (See figure 6)



Figure 6: A milk bottle being raised by multiple pulleys

What did you find?

- Each time the rope is looped over the rod, the force required to lift the bottle becomes less.
- This is an example of a block and tackle system of multiple pulleys.
- Even though proper pulleys are not used, the mechanical advantage of increasing the length of the rope and having the rope support the object being raised multiple times demonstrates the mechanical advantage of the system.
- Using real pulleys would decrease the friction of the system and therefore make the process even easier.

Note

Watch this video of an experiment to demonstrate the mechanical advantage of multiple pulleys.

Pulley'ing Your Own Weight (Duration: 01.32)



Summary

In this unit you have learnt the following:

- A pulley is a simple machine consisting of a grooved wheel in which a pulled rope or chain runs.
- A pulley changes the direction of a force to lift a load.
- Pulleys are divided into two types: a fixed pulley is attached to a non-moving point and the rope is attached to the object, while a movable pulley is attached to the object.
- Using a single pulley does not give any mechanical advantage, however using several pulleys together can create a mechanical advantage.
- The mechanical advantage of a system of pulleys reduces the effort by increasing the length of the rope used in the system.
- The equation to calculate the mechanical advantage of a pulley system is: MA = number of load supporting ropes
- Multiple pulley systems, called a block and tackle, are used to reduce the effort required to hoist very heavy objects.

Unit 2: Assessment

Suggested time to complete: 20 minutes

- 1. Define a pulley
- 2. Name the two kinds of pulleys and explain the difference between them.
- 3. Explain how to find the mechanical advantage of a pulley system.
- 4. What is the trade-off of mechanical advantage in a pulley system?
- 5. What are some examples of modern items that engineers have designed with pulleys?
- 6. A rock of 100 kg needs to be raised.
 - a. How much force would be required if a single pulley was used?
 - b. How much force would be required if a pulley system with five supporting ropes was used?
- 7. The engine of a vehicle needs to be raised by a block and tackle with four supporting chains. The force used is 500 N. Calculate the mass of the engine.
- 8. A single pulley does not provide a mechanical advantage (MA = 1).
 - a. Explain this statement.
 - b. What would be the benefit of using a single pulley to raise a heavy object?

The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Exercise 2.1

- 1. Possible answers: Crane, flagpole, blinds or shades, gym equipment, clothesline, water well, zip line, bike chain, elevator, sailboats.
- 2. To change the direction of force or make it easier to lift something.

Exercise 2.2

- 1. MA = number of load supporting ropes = 1
- 2. MA = number of load supporting ropes = 3
- 3. MA = number of load supporting ropes = 4

Back to Exercise 2.2

Unit 2: Assessment

- 1. A pulley is a simple machine consisting of a grooved wheel in which a pulled rope or chain runs.
- 2. A fixed pulley is attached to a fixed point and the rope is attached to the object. A movable pulley is attached to the object with one end of the rope attached to a fixed point and the other end of the rope is free.
- 3. The mechanical advantage of a pulley is system is equal to the number of ropes supporting the object being lifted or hauled.
- 4. The greater the mechanical advantage the greater the length of rope required.
- 5. Possible answers: Cranes, elevators, block and tackle on boats, flagpoles, zip lines, motors, bicycle rings/ chains, rock climbing devices, window blinds and sail boats.

6.

a.

Force required to lift 100 kg = $m \ge g = 100 \ge 9.8 = 980$ N MA = no. of supporting ropes = 1 980 N must be force applied b. Force required to lift 100 kg = $m \ge g = 100 \ge 9.8 = 980$ N

7.

MA = no. of supporting ropes = 4lifting force = 4 x applied force = 4 x 500 = 2 000 N $m = \frac{F}{g} = \frac{2\ 000}{9.8} = 204.08 \text{ kg}$

8.

a. The force applied (effort) is the same as the force required to raise the object (load).

MA =no. of supporting ropes = 5 980 ÷ 5 = 196 N must be force applied

b. A pulley can redirect the force so instead of pulling upwards, you can pull downwards which is easier to do because it is with gravity rather than against gravity.

Back to Unit 2: Assessment

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Unit 3: Gears and mechanical advantage

LEIGH KLEYNHANS



What you should know

Before you start this unit, make sure you can:

- Identify, describe, and apply principles of simple machines and mechanical advantage. See <u>level 2</u> subject outcome 2.4 to revise this.
- Define torque. See level 4 subject outcome 2.4, unit 1 to revise this.

Introduction¹

In this section you will see how gears can be used in a simple machine. A gear is a rotating machine part with one or more wheels with teeth usually made of metal or plastic. These wheels are called cogs.



Figure 1: A cog

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Gears are all around us and are found in just about everything that has spinning parts. For example, car engines, wind-up toys, non-digital clocks, drills, bicycles, car transmissions, powered wheelchairs, and lifts. Gears are used in machines to:

- change direction of rotation
- change speed of rotation
- change torque (force).

How do gears work?

Gears are arranged with two cogs that have their teeth locked into each other. The teeth in each cog are the same size. Since the teeth fit together, when one cog is turned, it causes the other cog to turn as well. If the first cog is rotated clockwise, the direction of the rotation of the other cog will be anti-clockwise (and vice versa). In this way gears can be used to change the direction of rotation.



Figure 2: A gear consisting of two cogs

When two cogs of different diameters are connected in a gear, slow rotation of the large cog will cause the smaller cog to rotate quickly. The gear in this case can increase the input speed of rotation to a higher output speed.

Turning a large cog slowly requires less energy than it would to rotate the small one quickly. By using cogs, there is saving of energy, making work easier. In this way, gears can change the torque and create mechanical advantage.

If cogs are different diameters, they will have different numbers of teeth. The ratio of the number of teeth in two meshed cogs is referred to as the gear ratio. For example, if one gear has 60 teeth and another has 20, the gear ratio when these two gears are connected together is 3:1 The smaller cog will complete three rotations for every one rotation of the bigger cog.

Note

Watch the following video which explains the basics of gears.

Gear Basics (Duration: 10.58)



Why are gears needed?

Some of the machines used in our daily life would be impossible to operate without gears. A car is a perfect example. The combustion and electrical engines that power cars use suitable gear systems to regulate engine power. For example, in a car transmission system, the driver (or an automated transmission mechanism) selects the appropriate gear ratio depending on the power required. On a flat road, a high gear ratio allows the vehicle to speed up, while when climbing up a hill, a small gear ratio produces the necessary (torque) force to move the vehicle up the slope.

Gear trains

Multiple gears are often connected together in gear trains, as illustrated in figure 3. When gears in a gear train are the same size, they rotate at the same speed and transfer the same torque or force. However, when the gears are different sizes either more torque/force or more speed will result.



Figure 3: A gear train

When we have two or more gears meshed together, the gear attached to the motor is called the input gear and the gear attached to the wheel is called the output gear. Any cogs between the input and the output are referred to as idle gears. The idle gear can ensure that the output and the input gear rotate in the same direction.

The gear ratio will be the ratio of the number of teeth on the output gear versus the number of teeth on the input gear (the drive gear). For the gear train in figure 3, the gear ratio will be 44 : 10 or 22 : 5.

When the input gear is larger than the output gear, the output gear rotates faster; this is called gearing up. The output gear provides less torque but moves at a higher speed. For example, many bicycles have different gears. When you would like the bike to go faster, you can change to a higher gear, so that it can provide speed, though it will do so with less force. When the input gear is smaller than the output gear, the output will rotate slower; this is called gearing down. Though it rotates slower, the torque from the output gear is more. In the example of a bicycle, when you want to climb a hill, you can change to a lower gear to get more force, though at a slower speed.

The difference between gears and pulleys

Now that you have studied both gears and pulleys, you can identify the similarities and differences between these two types of simple machines. Pulleys are mainly used to lift large, heavy objects providing a mechanical advantage so that less force is required. Gears are mainly used to increase speed of rotation with less torque.

However, pulleys can also be arranged to provide the same advantage as gears. The wheel of one pulley is driven by a motor and a belt is looped around this wheel and around a second wheel. When the motor is turned on, the powered wheel turns the belt which turns the second wheel.



Figure 4: A drive pulley system

If the two pulleys have different diameters, there will be a difference in the rotational speed. A pulley allows you to have more distance between the wheel axles, but they transmit less torque compared to gears, as they can allow slip to happen when the tension force exceeds the friction. With gears there is no slip, but the gear teeth need to be in contact with each other. Gear systems require a much more precise alignment than pulley systems.

Summary

In this unit you have learnt the following:

- A gear is a rotating machine part with cut teeth that mesh with another toothed part in order to transmit torque, increase speed or change direction of rotation.
- A gear train is a system of multiple gears.
- The gear ratio is the number of teeth on the input gear to the number of teeth on the output gear.
- Pulleys can be arranged to perform a similar function to gears.

Unit 3: Assessment

Suggested time to complete: 15 minutes

- 1. Define the following terms:
 - a. gear
 - b. gear train
 - c. gear ratio
- 2. How is a gear ratio designed for a particular application?
- 3. Give the gear ratios illustrated below and state the direction of rotation of the output gear:
 - a.



b.



- 4. In the gear assembly diagrams below, which one:
 - a. produces torque?
 - b. produces speed?
 - c. is neutral?



- 5. For each of the following situations, say whether you would use a pulley system, a gear system or either:
 - a. raising a flag
 - b. in a motorised garage door
 - c. lowering a heavy load onto the deck of a ship
 - d. rotating the drum of a washing machine

The <u>full solutions</u> are at the end of the unit.

Unit 3: Solutions

Unit 3: Assessment

1.

- a. a rotating machine part with cut teeth that mesh with another toothed part in order to transmit torque
- b. two or more gears connected together
- c. the ratio of the number of teeth on the input gear to the number of teeth on the output gear
- 2. Some applications (machines) may need to be fast, while other machines may need to be strong. When designing gears for a machine, it is important to know whether speed or torque (strength) is more important. In general, if you need to have more torque, select a large gear ratio. If you need to move fast, then use a smaller gear ratio.

3.

a. 1:1 clockwise

- b. 11:5 clockwise
- c. 1:1 clockwise

4.

- a. B output gear has larger diameter than input gear
- b. A output gear has smaller diameter than input gear
- c. C output and input gears are the same size

5.

- a. pulley
- b. either
- c. pulley
- d. either

Back to Unit 3: Assessment

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SUBJECT OUTCOME V WAVES, SOUND AND LIGHT: DESCRIBE AND APPLY PROPERTIES OF WAVES IN EVERYDAY LIFE CONTEXTS



Subject outcome

Subject outcome 3.1: Describe and apply properties of waves in everyday life contexts



Learning outcomes

- Identify and describe the basic concepts of the Doppler effect with sound and ultrasound.
- Apply the Doppler effect to radar and examples of sonic boom.
- Define EM radiation; identify the EM spectrum and identify examples (light, radio waves, microwave, IR, UV and others) and their application.
- Describe the dual nature of EM radiation; i.e. wave and particle nature.
- Identify the relationship of wavelength to frequency to explain colour and energy of light.
- Define lasers and give examples of its application.
- Explain and apply the photoelectric effect give examples of its application.



Unit 1 outcomes

By the end of this unit you will be able to:

- Identify and describe the basic concepts of the Doppler effect with sound and ultrasound.
- Apply the Doppler effect to radar and examples of sonic boom.



By the end of this unit you will be able to:

- Define electromagnetic (EM) radiation.
- Identify the EM spectrum and identify examples (light, radio waves, microwave, infrared, ultraviolet, x-rays and gamma rays).

- Describe the dual nature of EM radiation, i.e. wave and particle nature.
- Identify the relationship of wavelength to frequency to explain colour and energy of light.



Unit 3 outcomes

By the end of this unit you will be able to:

- Identify the applications of radiation across the EM spectrum.
- Define lasers and give examples of their application.
- \cdot Explain and apply the photoelectric effect and give examples of its application.

Unit 1: The Doppler effect

DYLAN BUSA



Apply the Doppler effect to radar and examples of sonic boom.

What you should know

Before you start this unit, make sure you can:

• Describe how the frequency of a sound wave affects the pitch of the sound. See <u>level 3 subject</u> <u>outcome 3.1 unit 2</u> if you need help with this.

Introduction

Even if you don't realise it, you have more than likely heard of the Doppler effect in action. Examples of it are the way ambulance and police car sirens change in pitch as they approach you or get further away.

Click <u>here</u> to listen to the Doppler effect.



Because of your experience, you instinctively know that the hooter's pitch going up indicates that the car is approaching and as the hooter's pitch decreases, the car is travelling away again. The Doppler effect does not only apply to sound waves though. It applies to all waves including light waves and other forms of electromagnetic radiation. To understand why, we need to take a closer look at what is happening.

Did you know?

The Doppler effect is named after the Austrian physicist Christian Doppler (see figure 1), who described the phenomenon in 1842. Doppler was a mathematician and physicist.



Figure 1: Christian Doppler

The Doppler effect

You may have noticed that the Doppler effect only happens when something is moving. Most of our everyday experiences of the Doppler effect occur when a vehicle, the source of the sound, is moving relative to us, the **observer** (or listener in the case of sound waves). In this case, the source is moving, and the observer is stationary.

Let's explore this simple case of the source moving relative to a stationary observer. As the source of the sound, a car's hooter for example, moves towards us, it is travelling in the same direction as the sound waves it emits that will reach our ears (see figure 2). But because it is moving towards us, the waves in front of the source get bunched up. This means that more waves pass a given point each second which means that there is an increase in the frequency. Because there is an increase in frequency, we hear a higher pitched sound.



Figure 2: The Doppler effect when a source is travelling towards an observer

It is important to note that the frequency of the sound wave from the source **does not change**. It is our **perception of this frequency** that has changed.

When a source moves away from us, the reverse happens. The waves behind the source get more spread out, resulting in a lower frequency and hence a lower perceived pitch (see figure 3).



Figure 3: The Doppler effect when a source is travelling away from an observer

However, it need not be the source that is moving. An observer who is moving relative to a stationary source will also experience the Doppler effect. An observer moving towards a source will experience the waves at a higher frequency and so will hear a higher pitched sound. An observer moving away from a source will experience the waves at a lower frequency and so will hear a lower pitched sound.

In fact, both the source and the observer can be moving. What is important is that there is **relative motion** between the source and the observer.



Time required: 10 minutes

What you need:

- \cdot a mobile phone or other small, light source of sound
- string (three to four metres long)

What to do:

1. Tie a three to four metre piece of string around your mobile phone (or some other small and light sound source). Make sure that the string is tied tightly and across all four sides like a bow around a present.



- 2. Go outside into a large, open space and play a song or some other sound on the mobile phone.
- 3. Ask a friend to stand nearby while you swing the phone carefully above your head making sure that you don't let it crash into anything or anyone.
- 4. Ask your friend to describe how the sound changes as the phone comes towards them and then moves away again.

- 5. Swop roles so that you can hear the sound effect of the phone moving towards and away from you.
- 6. Does changing the speed of the phone affect how much the sound changes?

What did you find?

As the phone moved through the air you should have heard a change in the pitch of the sound coming from the phone. As the phone moved towards you, the pitch of the sound goes up. As the phone moves away from you, the pitch of the sound goes down.

This is because, as the phone moves towards you, the sound waves in front of the phone get bunched up and so more waves pass a certain point in a given time. There is a perceived change in the frequency of the sound and so the pitch of the sound increases.

The exact opposite happens as the phone moves away from you.

You should also have found that the extent of the change in the pitch of the sound coming from the phone (both the increase and decrease in pitch) was greater the faster the phone was moving. The faster the phone moves, the more bunched up or spread out the sound waves become and so the greater the extent in the perceived change in frequency and pitch.

Note

If you have an internet connection, watch the video called What is the Doppler Effect? for an excellent explanation and demonstration of the Doppler effect.

What is the Doppler Effect? (Duration: 02.30)



You should also play around with the interactive Doppler effect simulation at this link.



Change the velocity of the source (v_S) to see how this affects the extent of the Doppler effect experienced by the Observer (R). Notice that the frequency of the waves emitted by the source never changes. What changes is the perceived frequency of the waves by the observer. Each time a wave front hits the observer, it blinks.

Calculate the Doppler effect

If we know the speed of the waves, the frequency of the waves and the relative speeds of the source and observer, we can calculate the frequency that the observer will perceive using the following formula.



Remember that v_L and v_S are vectors. The signs of these quantities will tell us if the relative motion of the source and observer (or listener) is towards or away from each other.

Take note!

If the source is moving towards the observer, vs is negative.
If the observer is moving towards the source, vL is positive.
Example 1.1

A police car with a siren sounding at 750 Hz is driving along a road at 22 m.s^{-1} . You are standing on the pavement. If the speed of sound in air is 344 m.s^{-1} ., what frequency will you hear as:

- 1. the car travels towards you?
- 2. the car travels away from you?

Solutions

All the values given are in SI units so we do not need to do any conversions before we can calculate. We have been asked in each case for the frequency perceived by the listener.

1.

$$egin{aligned} f_S &= 750 \; \mathrm{Hz} \ v &= 344 \; \mathrm{m.s^{-1}} \ v_L &= 0 \; \mathrm{m.s^{-1}} & \mathrm{listener} \ v_S &= -22 \; \mathrm{m.s^{-1}} & \mathrm{source} \; \mathrm{s} \ f_L &= \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ \therefore f_L &= \left(rac{344 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}}{344 \; \mathrm{m.s^{-1}} - 22 \; \mathrm{m.s^{-1}}}
ight) \cdot 750 \; \mathrm{Hz} \ \therefore f_L &= 801 \; 24 \; \mathrm{Hz} \end{aligned}$$

is stationary is moving towards listener

 $\therefore f_L = 801.24 \text{ Hz}$

This answer makes sense as we expect the observer to perceive the siren at a higher pitch (i.e. a higher frequency) as the car comes towards them.

2.

 $f_S = 750 \text{ Hz}$ $v=344~\mathrm{m.s}^{-1}$ $v_L = 0 \,\,\mathrm{m.s^{-1}}$ listener is stationary $v_S=22\,\,\mathrm{m.s}^{-1}$ source is moving away from the listener $f_L = \left(rac{v\pm v_L}{v\pm v_S}
ight)\cdot f_S$ $\therefore f_L = \left(rac{344 ext{ m.s}^{-1} + 0 ext{ m.s}^{-1}}{344 ext{ m.s}^{-1} + 22 ext{ m.s}^{-1}}
ight) \cdot 750 ext{ Hz}$ $\therefore f_L = 704.92 \; \mathrm{Hz}$

This answer makes sense as we expect the observer to perceive the siren at a lower pitch (i.e. a lower frequency) as the car moves away from them.

Example 1.2

A fire truck is driving towards a building in which a fire alarm is ringing at 825 Hz. The truck is travelling at $_{18\ m.s^{-1}}$. If the speed of sound in air is $_{344\ m.s^{-1}}$, what frequency do the firemen in the truck hear?

Solution

All the values given are in SI units so we do not need to do any conversions before we can calculate. We have been asked for the frequency perceived by the listener.

$$egin{aligned} &f_S = 825 \ {
m Hz} \ &v = 344 \ {
m m.s}^{-1} \ &v_L = 18 \ {
m m.s}^{-1} \ &v_S = 0 \ {
m m.s}^{-1} \ &f_L = \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ &\therefore f_L = \left(rac{344 \ {
m m.s}^{-1} + 18 \ {
m m.s}^{-1} }{344 \ {
m m.s}^{-1} + 0 \ {
m m.s}^{-1}}
ight) \cdot 825 \ {
m Hz} \ &\therefore f_L = 868.17 \ {
m Hz} \end{aligned}$$

listener is moving towards the source source is stationary

This answer makes sense as we expect the observer to perceive the siren at a higher pitch (i.e. a higher frequency) as they move towards the source.



Example 1.3

A hearing-impaired passenger is at a train station. They can only hear frequencies above 900 Hz. The train approaches the platform at a constant speed of $85~{
m km/h}$ and sounds its whistle as a warning. The whistle sounds at 815 Hz. Assume the speed of sound in air is 344 m.s^{-1} .

- 1. Will the person hear the warning whistle? Explain your answer.
- 2. What frequency would the whistle need to be for the person to hear it if the train travelled at the same speed of 85 km/h?
- 3. At what speed would the train need to travel for the person to hear the whistle if the frequency is 800 Hz?

Solution

Not all the values given are in SI units. We need to convert the speed of the train to ${
m m.s^{-1}}$. $85 \text{ km/h} = 23.61 \text{ m.s}^{-1}$.

1. We have been asked for the frequency perceived by the listener.

$$egin{aligned} f_S &= 815 \; \mathrm{Hz} \ v &= 344 \; \mathrm{m.s^{-1}} \ v_L &= 0 \; \mathrm{m.s^{-1}} \ v_S &= -23.61 \; \mathrm{m.s^{-1}} \ f_L &= \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ dots f_L &= \left(rac{344 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}} \ 344 \; \mathrm{m.s^{-1}} - 23.61 \; \mathrm{m.s^{-1}}
ight) \cdot 815 \; \mathrm{Hz} \ dots f_L &= 875.06 \; \mathrm{Hz} \end{aligned}$$

listener is stationary source is moving towards listener

 $\therefore f_L = 875.06 \text{ Hz}$

The frequency of the sound of the train whistle perceived by the observer is $875.06~\mathrm{Hz}$. This means that the person will not be able to hear the sound as it is below their audible threshold.

listener is stationary

source is moving towards listener

2. We have been asked to calculate the frequency of the whistle such that the frequency perceived by the listener would be 900 Hz. $f_{r} = 900 \text{ Hz}$

$$\begin{split} y_L &= 300 \text{ mz} \\ v &= 344 \text{ m.s}^{-1} \\ v_L &= 0 \text{ m.s}^{-1} \\ v_S &= -23.61 \text{ m.s}^{-1} \\ f_L &= \left(\frac{v \pm v_L}{v \pm v_S}\right) \cdot f_S \\ \therefore f_S &= \frac{f_L}{\left(\frac{v \pm v_L}{v \pm v_S}\right)} \\ \therefore f_S &= \frac{900}{\left(\frac{344 \text{ m.s}^{-1} \pm 0 \text{ m.s}^{-1}}{344 \text{ m.s}^{-1} - 23.61 \text{ m.s}^{-1}}\right)} \end{split}$$

 $\therefore f_S = 838.23 \text{ Hz}$

The frequency of the sound of the train whistle would need to be at least $838.23~\mathrm{Hz}$ to be heard by the listener.

3. We have been asked to calculate the speed that the train must travel at such that the frequency perceived by the listener would be 900 Hz of a whistle with a frequency of 815 Hz.

$$\begin{split} f_L &= 900 \text{ Hz} \\ f_S &= 815 \text{ Hz} \\ v &= 344 \text{ m.s}^{-1} \\ v_L &= 0 \text{ m.s}^{-1} \\ f_L &= \left(\frac{v \pm v_L}{v \pm v_S}\right) \cdot f_S \\ \therefore \frac{f_S}{f_L} &= \frac{v \pm v_S}{v \pm v_L} \\ \therefore \frac{f_S}{f_L} \cdot (v \pm v_L) &= v \pm v_S \\ \therefore v_S &= \frac{f_S}{f_L} \cdot (v \pm v_L) - v \\ \therefore v_S &= \frac{815 \text{ Hz}}{900 \text{ Hz}} \times (344 \text{ m.s}^{-1} + 0 \text{ m.s}^{-1}) - 344 \text{ m.s}^{-1} \\ &= -32.49 \text{ m.s}^{-1} \end{split}$$

The speed of the approaching train would need to be at least 32.49 m.s^{-1} towards the listener for the train whistle to be heard by the listener.

Note

You will only ever need to answer questions where either the source or the observer are moving, never both.



Exercise 1.1

1. Passengers in a car hear its hooter at a frequency of 812 Hz. What frequency does someone standing on the pavement hear as the car moves directly towards them at a speed of 85 km/h? Assume the speed of sound in air is 340 m.s^{-1} .

Question 2 adapted from Everything Science Grade 12 Exercise 6-1 question 3

- 2. A small plane is taxiing directly away from you down a runway. The noise of the engine, as the pilot hears it, has a frequency 1.15 times the frequency that you hear. What is the speed of the plane? Assume the speed of sound in air is 340 m.s^{-1} .
- 3. An experiment was done in the South Pole. The temperature was -40° C. A source, moving away from an observer at 22.5 m.s^{-1} , emitted a sound at 2500 Hz. The observer measured a frequency of only 2335 Hz. What was the speed of sound at this cold temperature?

Question 4 adapted from Everything Science Grade 12 Exercise 6-1 question 5

4. Cecil approaches a source emitting a sound with a frequency of $437.1~{
m Hz}$. Assume the speed of sound is $344~{
m m.s}^{-1}$.

- a. How fast does Cecil need to move to observe a frequency that is 20 percent higher?
- b. If he passes the source at this speed, what frequency will he measure when he is moving away?

The <u>full solutions</u> are at the end of the unit.

Applications of the Doppler effect

As we said above, the Doppler effect applies to all waves. Light from a source travelling towards you will be perceived to have a slightly higher frequency and will shift blue (blueshift). Light from a source travelling away from you will be perceived to have a slightly lower frequency and will shift red (redshift). This is illustrated in figure 4.



Figure 4: Redshift and blueshiftThis effect is used to determine whether distant galaxies are travelling towards or away from us. The extent of the blueshift or redshift can be used to determine how fast another galaxy is approaching or receding.

Traffic cops use exactly the same technique to measure the speed of motorists using radio waves. This is called **radar**.

The radar gun (see figure 5) fires a burst of radio waves towards an approaching vehicle. If the vehicle was stationary, these radio waves would be reflected back at exactly the same frequency. However, if the car is moving towards the gun, the waves get reflected back at a slightly higher frequency. The gun compares the frequency of the reflected waves with the source waves and uses the difference to calculate the speed of the vehicle. The faster the vehicle is travelling, the greater the change in frequency.

The same principle underpins all radar applications including those used to track the path of aircraft as well as storms.


Sonic booms

The phenomena of sonic booms can be thought of as the Doppler effect on steroids. Have a look at the formula we have been using to calculate the frequency of the sound perceived by the observer.

$$f_L = \left(rac{v\pm v_L}{v\pm v_S}
ight)\cdot f_S$$

What happens as v_S gets closer and closer to v? Can you see that the denominator gets very, very small meaning that the whole fraction gets very, very big. This means that f_L gets very big and approaches infinity as v_S approaches v.

When this happens, a sonic boom occurs. A sonic boom is the sound associated with the shock waves created by an object travelling through the air, at or faster than, the speed of sound. They create a large amount of energy, and sound like explosions. The 'crack' of a bullwhip is a result of a sonic boom.



As the speed of the sound source increases to the speed of the sound waves themselves, the waves, not being able to get out of each other's way, are forced together. They eventually merge, through constructive interference, into a single shock wave traveling at the speed of sound. This happens at the critical speed known as **Mach**. The shock waves radiate out from the sound source in the shape of a cone (see figure 6). When a person on the ground hears the sonic boom, the object that made it has long since passed.



Figure 6: A sonic boom cone

Did you know?

Chuck Yeager of the United States Airforce was the first person to break the sound barrier. He flew a small airplane called the Bell X-1 (see figure 7) which is currently in the Smithsonian Air and Space Museum in Washington, DC.



Summary

In this unit you have learnt the following:

- When a source emitting sound relative to a listener, the listener perceives a change in the frequency of the sound. This is called the Doppler effect.
- If the source moves relatively towards the listener, the perceived frequency is higher.
- If the source moves relatively away from the listener, the perceived frequency is lower.
- The Doppler effect also occurs when a light emitting object moves relative to an observer.
- If the object is moving relatively towards the observer, the light appears more blue (blueshift to higher frequency light).
- If the object is moving relatively away from the observer, the light appears more red (redshift to lower frequency light).
- \cdot The frequency of the wave (sound or light) perceived by the observer can be calculated using the

formula
$$f_L = \left(rac{v\pm v_L}{v\pm v_S}
ight)\cdot f_S.$$

Unit 1: Assessment

Suggested time to complete: 35 minutes

Question 1 adapted from NC(V) Level 4 Physical Science Paper 1 February 2019 question 10

- 1. A goods train sounds its horn as it approaches and passes through a station at a constant speed of $20 \text{ m} \cdot \text{s}^{-1}$. The horn has a frequency of 550 Hz. Assume the speed of sound in air is $344 \text{ m} \cdot \text{s}^{-1}$.
 - a. What is the frequency of the horn heard by people sitting on the station platform as the train approaches?
 - b. What is the frequency of the horn heard by the driver of the train?
 - c. The train passes through the station. The driver next sounds the horn when the train is travelling away from the station and has accelerated to a constant speed of $_{30\ m.s}^{-1}$. Will the frequency of the horn heard by the people sitting on the station platform INCREASE/DECREASE/REMAIN THE SAME? Support your answer with a calculation.

Question 2 adapted from NC(V) Level 4 Physical Science Paper 1 November 2019 question 10

- 2. The siren of a police car emits sound waves of wavelength 0.55 m. A stationary listener watches the police car approaching him at a constant velocity on a straight road. Assume that the speed of sound in air is 345 m.s^{-1} .
 - a. How does the wavelength of the sound waves change as the police car approaches the listener? Only write down INCREASES, DECREASES or REMAINS THE SAME.
 - b. Calculate the observed frequency of the sound waves if the car moves toward him at a speed of $120 \ \rm km/h?$
 - c. How will the answer to question 2.b. change if the police car moves away from the listener at 120 km/h? Only write down INCREASES, DECREASES or REMAINS THE SAME.
- 3. A motorbike approaches you at an unknown speed. The sound of the bike's engine has a frequency of 510 Hz, however you hear a frequency of 545 Hz. The speed of sound is 344 m.s^{-1} .
 - a. Calculate the speed of the motorbike.
 - b. How will the sound change as the bike passes you? Explain this phenomenon in terms of the wavelength and frequency of the sound.

Question 4 adapted from Everything Science Grade 12 Exercise 6-2 question 4

4. A police car is driving towards a fleeing suspect at $\frac{v}{35}$ m.s⁻¹, where v is the speed of sound. The frequency of the police car's siren is 400 Hz. The suspect is running away at $\frac{v}{68}$ m.s⁻¹. What frequency does the suspect hear?

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1. 85 km/h = 23.61 ${\rm m.s^{-1}}$ We have been asked to calculate f_L .

$$egin{aligned} & f_S = 812 \; \mathrm{Hz} \ & v = 340 \; \mathrm{m.s^{-1}} \ & v_L = 0 \; \mathrm{m.s^{-1}} \ & v_L = 0 \; \mathrm{m.s^{-1}} \ & v_S = -23.61 \; \mathrm{m.s^{-1}} \ & source \; \mathrm{is \; stationary} \ & v_S = -23.61 \; \mathrm{m.s^{-1}} \ & f_L = \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ & f_L = \left(rac{340 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}}{340 \; \mathrm{m.s^{-1}} - 23.61 \; \mathrm{m.s^{-1}}}
ight) \cdot 812 \; \mathrm{Hz} \ & = 872.59 \; \mathrm{Hz} \end{aligned}$$

2. We have been asked to calculate v_S .

$$egin{aligned} &f_L = x \; \mathrm{Hz} \ &f_S = 1.15x \; \mathrm{Hz} \ &v = 340 \; \mathrm{m.s^{-1}} \ &v_L = 0 \; \mathrm{m.s^{-1}} \ &v_L = \left(\frac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ &\therefore v_S = \frac{f_S}{f_L} \cdot (v \pm v_L) - v \ &\therefore v_S = \frac{f_S}{f_L} \cdot (v \pm v_L) - v \ &\therefore v_S = \frac{1.15x}{x} \cdot (340 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}) - 340 \; \mathrm{m.s^{-1}} \ &\therefore v_S = 1.15 \cdot (340 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}) - 340 \; \mathrm{m.s^{-1}} \ &= 51 \; \mathrm{m.s^{-1}} \end{aligned}$$

Therefore $v_S=51\,\,{
m m.s}^{-1}$ away from the observer.

3. We have been asked to calculate v.

$$\begin{array}{l} f_L = 2\ 335\ \mathrm{Hz} \\ f_S = 2\ 500\ \mathrm{Hz} \\ v_S = 22.5\ \mathrm{m.s^{-1}} \\ v_L = 0\ \mathrm{m.s^{-1}} \\ f_L = \left(\frac{v\pm v_L}{v\pm v_S}\right)\cdot f_S \\ \vdots \frac{f_L}{f_S} = \frac{v\pm v_L}{v\pm v_S} \\ \vdots \frac{v\pm v_L}{f_S} = \frac{2\ 335\ \mathrm{Hz}}{2\ 500\ \mathrm{Hz}} = 0.934 \\ \vdots v + 0\ \mathrm{m.s^{-1}} = 0.934 \times (v + 22.5\ \mathrm{m.s^{-1}}) \\ \vdots v = 0.934v + 21.015\ \mathrm{m.s^{-1}} \\ \vdots v(1 - 0.934) = 21.015\ \mathrm{m.s^{-1}} \\ \vdots v = 318.41\ \mathrm{m.s^{-1}} \end{array}$$

4.

a. We have been asked to calculate v_L .

$$egin{aligned} &f_L = 437.1 imes 1.2 \; \mathrm{Hz} = 524.52 \; \mathrm{Hz} \ &f_S = 437.1 \; \mathrm{Hz} \ &v = 344 \; \mathrm{m.s^{-1}} \ &v_S = 0 \; \mathrm{m.s^{-1}} \ &v_S = 0 \; \mathrm{m.s^{-1}} \ &t_L = \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ &\therefore v_L = rac{524.52 \; \mathrm{Hz}}{437.1 \; \mathrm{Hz}} \cdot (344 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}) - 344 \; \mathrm{m.s^{-1}} \ &= 68.8 \; \mathrm{m.s^{-1}} \end{aligned}$$

b. We have been asked to calculate f_L .

$$egin{aligned} &f_S = 437.1 \; \mathrm{Hz} \ &v = 344 \; \mathrm{m.s^{-1}} \ &v_S = 0 \; \mathrm{m.s^{-1}} \ &v_L = -68.8 \; \mathrm{m.s^{-1}} \ &f_L = \left(rac{v \pm v_L}{v \pm v_S}\right) \cdot f_S \ &\therefore f_L = \left(rac{344 \; \mathrm{m.s^{-1}} - 68.8 \; \mathrm{m.s^{-1}}}{344 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}}
ight) \cdot 437.1 \; \mathrm{Hz} \ &= 349.68 \; \mathrm{Hz} \end{aligned}$$

source is stationary observer moving away from source

Back to Exercise 1.1

Unit 1: Assessment

1.

a. We have been asked to calculate f_L . $f_S = 550 \text{ Hz}$ $v = 344 \text{ m.s}^{-1}$ $v_L = 0 \text{ m.s}^{-1}$ observer is stationary $v_S = -20 \text{ m.s}^{-1}$ source moving towards observer $f_L = \left(\frac{v \pm v_L}{v \pm v_S}\right) \cdot f_S$ $\therefore f_L = \left(\frac{344 \text{ m.s}^{-1} + 0 \text{ m.s}^{-1}}{344 \text{ m.s}^{-1} - 20 \text{ m.s}^{-1}}\right) \cdot 550 \text{ Hz}$ = 583.95 Hz

observer is stationary

source moving away from observer

- b. The driver will hear the actual frequency of the horn $550\,\,\mathrm{Hz}$
- c. The frequency of the horn for the people will decrease.

$$egin{aligned} f_S &= 550 \; \mathrm{Hz} \ v &= 344 \; \mathrm{m.s^{-1}} \ v_L &= 0 \; \mathrm{m.s^{-1}} \ v_S &= 30 \; \mathrm{m.s^{-1}} \ f_L &= \left(rac{v \pm v_L}{v \pm v_S}
ight) \cdot f_S \ \therefore f_L &= \left(rac{344 \; \mathrm{m.s^{-1}} + 0 \; \mathrm{m.s^{-1}}}{344 \; \mathrm{m.s^{-1}} + 30 \; \mathrm{m.s^{-1}}}
ight) \cdot 550 \; \mathrm{Hz} \ &= 505.88 \; \mathrm{Hz} \end{aligned}$$

2.

- a. Increases
- b. $120 \ \mathrm{km/h} = 33.33 \ \mathrm{m.s^{-1}}$

 $v = f imes \lambda$ $\therefore f = rac{v}{\lambda} = rac{345 \text{ m.s}^{-1}}{0.55 \text{ m}} = 627.27 \text{ Hz}$ We have been asked to calculate f_L .

$$\begin{split} f_S &= 627.27 \text{ Hz} \\ v &= 345 \text{ m.s}^{-1} \\ v_L &= 0 \text{ m.s}^{-1} \\ v_S &= -33.33 \text{ m.s}^{-1} \\ f_L &= \left(\frac{v \pm v_L}{v \pm v_S}\right) \cdot f_S \\ \therefore f_L &= \left(\frac{345 \text{ m.s}^{-1} + 0 \text{ m.s}^{-1}}{345 \text{ m.s}^{-1} - 33.33 \text{ m.s}^{-1}}\right) \cdot 627.27 \text{ Hz} \\ &= 694.35 \text{ Hz} \\ \end{split}$$
We have been asked to calculate v_S .

$$\begin{aligned} f_S &= 510 \text{ Hz} \\ v_L &= 575 \text{ Hz} \\ v &= 344 \text{ m.s}^{-1} \\ v_L &= 0 \text{ m.s}^{-1} \\ v_L &= \frac{f_S}{f_L} \cdot (v \pm v_L) - v \\ \therefore v_S &= \frac{f_S}{f_L} \cdot (v \pm v_L) - v \\ \therefore v_S &= \frac{510 \text{ Hz}}{575 \text{ Hz}} \cdot (344 \text{ m.s}^{-1} + 0 \text{ m.s}^{-1}) - 344 \text{ m.s}^{-1} \\ \end{split}$$

$$575 \text{ Hz}$$

= 38.89 m.s⁻¹

C.

a.

3.

b. As the bike passes and travels away, the sound waves will become more spread out. The listener will perceive a sound of a lower frequency and hence a sound with a lower pitch.

4. We have been asked to calculate f_L . Both the car and the suspect are moving. It is clear that the car is moving towards the suspect. However, because the car is moving faster than the suspect, the suspect is also moving towards the car even though he is trying to run away.

$$\begin{split} f_{S} &= 400 \text{ Hz} \\ v &= v \text{ m.s}^{-1} \\ v_{L} &= \frac{v}{68} \text{ m.s}^{-1} \\ v_{S} &= -\frac{v}{35} \text{ m.s}^{-1} \\ f_{L} &= \left(\frac{v \pm v_{L}}{v \pm v_{S}}\right) \cdot f_{S} \\ &= \left(\frac{v + \frac{v}{68}}{v - \frac{v}{35}}\right) \cdot 400 \text{ Hz} \\ &= \left(\frac{\frac{68v + v}{68}}{\frac{35v - v}{35}}\right) \cdot 400 \text{ Hz} \\ &= \left(\frac{68v + v}{68} \times \frac{35v - v}{35}\right) \cdot 400 \text{ Hz} \\ &= \left(\frac{2380v + 35v}{2380v - 60v}\right) \cdot 400 \text{ Hz} \\ &= \left(\frac{v(2380 + 35)}{v(2380 - 60)}\right) \cdot 400 \text{ Hz} \end{split}$$

 $= 416.38~\mathrm{Hz}$

observer is moving towards source

source is moving towards observer

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Unit 2: Electromagnetic radiation

DYLAN BUSA



Unit outcomes

By the end of this unit you will be able to:

- Define electromagnetic (EM) radiation.
- Identify the EM spectrum and identify examples (light, radio waves, microwave, infrared, ultraviolet, x-rays and gamma rays).
- $\cdot\;$ Describe the dual nature of EM radiation, i.e. wave and particle nature.
- Identify the relationship of wavelength to frequency to explain colour and energy of light.

What you should know

Before you start this unit, make sure you can:

- Explain how a wave is a periodic disturbance in a medium that transfers energy from one point to another. Refer to <u>level 2 subject outcome 3.1 unit 1</u> if you need help with this.
- Explain the differences and similarities between transverse and longitudinal waves. Refer to <u>level 2</u> subject outcome 3.1 unit 2 if you need help with this.
- Explain what is meant by wavelength, frequency, amplitude and period. Refer to <u>level 2 subject</u> <u>outcome 3.1 unit</u> 3 if you need help with this.
- Describe the electric field that exists around a point charge. Refer to <u>level 3 subject outcome 4.1 unit 1</u> if you need help with this.
- Explain the principles of electromagnetism, especially how a moving charge produces a magnetic field and how a changing magnetic field produces or applies a force on a charged particle. Refer to <u>level 3</u> subject outcome 4.2 units 1 and 2 if you need help with this.

Introduction

Sight is one of the most amazing senses we have. It allows us to interact with and experience the world in wonderful ways. Think of the beautiful sunsets you might have seen (see figure 1). So much of life is mediated through the gift of sight.



Figure 1: A beautiful sunset

But despite how amazing our sense of sight is, did you know that we can actually see very little of what there is to see with our eyes? Did you know that there are kinds of 'light waves' that you cannot see with your eyes? In fact, the light that we can see is just a small portion of a larger family of waves called the **electromagnetic spectrum**. We will see later on why it is called the electromagnetic spectrum.

The electromagnetic spectrum

The so-called electromagnetic spectrum (the EM spectrum) consists of an extremely wide range of wave frequencies that transmit energy in the form of electromagnetic radiation. On the one end of the spectrum, we have radio waves which can have wavelengths of more than $1\ 000\ {\rm km}$. On the other end are gamma waves that can have wavelengths smaller than the nucleus of an atom. Figure 2 shows a summary of all the different types of waves that make up the EM spectrum.

As we move from left to right in figure 2, the wavelength of the waves gets shorter and shorter. This means that the frequency of the waves gets higher and higher. The higher the frequency of the wave, the more energy it can carry. This is why ultraviolet radiation, x-rays and gamma rays can be so dangerous. This radiation has enough energy to alter some chemical bonds in our cells which can lead to serious illness.



Figure 2: The electromagnetic spectrum

Somewhere in the middle of the EM spectrum is a little band of waves with wavelengths between about 700 nm and 400 nm called visible light that our eyes can see (see figure 3). By the way, a nanometre (nm) is one billionth of a metre.



Figure 3: The visible light part of the EM spectrum

Figure 3 also shows us that red light has a longer wavelength (and hence lower frequency) than blue light. The higher frequency of blue light means that it carries more energy as well.



Tour of the Electromagnetic Spectrum - 01 - Introduction (Duration: 05.03)



This video is the first video in a series of videos that explore each part of the spectrum in detail. Why not watch the rest of the series?

• Tour of the Electromagnetic Spectrum – 02 – Radio Waves.



<u> Tour of the Electromagnetic Spectrum – 02 – Radio Waves</u> (Duration: 03.38

• Tour of the Electromagnetic Spectrum – 03 – Microwaves.



• Tour of the Electromagnetic Spectrum – 04 – Infrared Waves.



Tour of the Electromagnetic Spectrum – 04 – Infrared Waves (Duration: 05.23)

• Tour of the Electromagnetic Spectrum 05 – Visible Light Waves.





We will investigate each part of the EM spectrum and its applications in more detail in the next unit.

Properties of EM radiation

The waves in this spectrum have some important and unique properties. Firstly, you might remember in level 2 when we first talked about waves that we said that waves transfer energy through a medium and need a medium through which to travel. That is certainly true for water waves and sound waves, but it is not true for electromagnetic radiation waves. They do not need a medium and can propagate through empty space. They can, however, also travel through media such as air, water and you.



Secondly, we know that sound waves travel through the air at about 340 m.s^{-1} (depending on the temperature). Next to EM waves, they are almost standing still. EM waves travel through a vacuum at about

300 000 ${\rm m.s}^{-1}$. That is almost 900 000 times faster! In fact, EM waves travel at the speed limit of the universe which we call the speed of light. We give this constant the symbol *c*. The '*c*' in $E = mc^2$ is the speed of light. Just like other waves however, $v = f \times \lambda$ or for EM waves, $c = f \times \lambda$.

And lastly, EM waves are **dual-natured**. Depending on how we look at them they can behave like the ordinary transverse waves we have become familiar with, or they can behave like particles. We call the individual particles or packets of EM radiation **photons**.



Electromagnetic waves

You might think it strange to refer to light as an electromagnetic wave. What is electric or magnetic about light? This does not fit our everyday experience of light.

In <u>level 3 subject outcome 4.1 unit 1</u>, we drew the electric field that exists around a point charge. The charged particle creates an electric field that radiates out from the charge. Now, if we move this charged particle up and down, we will create a changing electric field in a particular region near the particle.

We know that a magnetic field is produced around a current-carrying wire. The reason is that there is a changing electric field around the wire. So, the changing electric field created by our oscillating charged particle will create a changing magnetic field.

Similarly, we know that a changing magnetic field results in a changing electric field. If there is a wire in the changing magnetic field, the electrons are made to move through the wire by the changing electric field. This is the basis of how generators work.

So, a changing electric field creates a changing magnetic field and a changing magnetic field creates an electric field. We have a positive feedback loop or chain reaction, and it is this mutual induction situation which propagates out as an electromagnetic wave due to the initial oscillation of our charged particle.

It is called an electromagnetic wave because it consists of an oscillating **electric field** and an oscillating **magnetic field**. These fields oscillate at right angles to each other and at right angles to the direction of the wave (see figure 4).

Electromagnetic Wave



Figure 4: An electromagnetic wave

Figure 5 is an animation showing how an EM wave propagates.



Figure 5: EM wave propagation

Note

If you have an internet connection, you should watch the video called Electromagnetic waves and the electromagnetic spectrum for a more detailed explanation of what electromagnetic waves are and how they are created.

Electromagnetic waves and the electromagnetic spectrum (Duration: 05.47)





Example 2.1

Calculate the frequency of an electromagnetic radio wave with a wavelength of $5.1\,\,\mathrm{m}$.

Solution

EM waves obey the same wave equation that we have used before. The only difference is that we replace 'v' in the equation with 'c', the speed of light. Therefore, $c = f \times \lambda$ where $c = 300 \ 000 \ \text{m.s}^{-1}$.

We have been asked to calculate f. All the given information is in SI units. $c = 300 \ 000 \ \mathrm{m.s^{-1}}$ (unless you are told otherwise, you can assume this value of c for the speed of light in a vacuum)

$$egin{aligned} \lambda &= 5.1 \ \mathrm{m} \ &c &= f imes \lambda \ dots &f &= rac{c}{\lambda} \ &= rac{300\ 000\ 000\ \mathrm{m.s^{-1}}}{5.1\ \mathrm{m}} \end{aligned}$$

$$= 58 823 529.41 \text{ Hz}$$

When we get very big answers like this, we usually convert the units or use scientific notation to simplify them. We can write the answer as 58.823 MHz or 5.8823×10^7 Hz.



Exercise 2.1

- 1. Red light has a wavelength of about 700 nm. Calculate the frequency of red light.
- 2. An x-ray machine produces radiation with a frequency of 2.1×10^{18} Hz. What is the wavelength of these waves?
- 3. A space telescope detects EM radiation with a frequency of 3.25×10^{12} Hz. Can the human eye detect this radiation? Show a calculation to support your answer.

The <u>full solutions</u> are at the end of the unit.

Wave-particle duality

The section above explains how electromagnetic radiation behaves like a wave. It is the result of changing electric and magnetic fields that propagate out from a source like a wave. Light displays the same interference patterns associated with other kinds of waves. Wave-like properties like wavelength and frequency neatly explain the different energies associated with different parts of the EM spectrum as well as properties of light like colour.

But sometimes, light does not behave like a wave at all. Sometimes it behaves like a beam of particles. For example, if you shine light onto a metal surface, the light transfers some of its energy to the atoms in the metal but only in very specific and discrete amounts called **quanta**. In this case, it is as though light was a beam of little packets of energy. We call these packets of energy **photons**.

The energy of each photon is related to the wavelength of the EM radiation by a special number called

Planck's constant, named after Max Planck. Planck's constant is given the symbol h and $h = 6.63 \times 10^{-34} \text{ J.Hz}^{-1}$. Sometimes you may see Planck's constant written as $h = 6.63 \times 10^{-34} \text{ J.s}$. Although this is technically correct J.s are not the official SI units.



The energy of a photon is calculated using E=hf or $E=h\cdot rac{c}{\lambda}$ where:

 ${\it E}$ is the energy of the photon in ${\rm J}$

 $h=6.63 imes 10^{-34}~
m J.Hz^{-1}$ (Planck's constant)

f is the frequency of the wave in $_{
m Hz}$

 $c=300\,\,000\,\,000\,\,\mathrm{m.s^{-1}}$ (the speed of light in a vacuum)

 λ is the wavelength of the wave in m

Did you know?

This duality – sometimes behaving like a wave and sometimes behaving like a particle – means that EM radiation is really like nothing else. It was the discovery of this dual nature of EM radiation that led to quantum mechanics.

Note

For more on the duality of EM radiation watch the video called "Is light a particle or a wave?".

Is light a particle or a wave? (Duration: 04.23)



Example 2.2

Calculate the energy of a photon with a frequency of $5.6 imes 10^{17}~{
m Hz}$.

Solution

We have been asked to calculate the energy of a photon and have been given the frequency. Therefore, we need to use the equation E = hf. All the values have been given in standard units.

$$E = hf$$

 $\therefore E = 6.63 imes 10^{-34} \, \mathrm{J.Hz^{-1}} imes 5.6 imes 10^{17} \, \mathrm{Hz}$
 $= 3.7128 imes 10^{-16} \, \mathrm{J}$

The energy of one photon is $3.7128 imes 10^{-16}$ J.

Example 2.3

Calculate the energy of one photon of infrared light with a wavelength of $0.021~\mathrm{mm}$.

Solution

We have been asked to calculate the energy of a photon and have been given the wavelength. Therefore, we need to use the equation $E = h \cdot \frac{c}{\lambda}$. The wavelength is not in standard units and needs to be converted.

$$0.021 \,\,\mathrm{mm} = 2.1 imes 10^{-5} \,\,\mathrm{m}$$

$$E = h \cdot rac{c}{\lambda}$$

 $\therefore E = 6.63 \cdot rac{10^{-34}}{2.1 imes 10^{-5}} \mathrm{m}$
 $= 9.47 imes 10^{-21} \mathrm{J}$

The energy of one photon is $9.47 imes 10^{-21}$ J.



Exercise 2.2

- 1. Calculate the energy of a photon of EM radiation with a frequency of $4.67 imes 10^{12}$ Hz.
- 2. Determine the energy of a photon of orange EM radiation with a wavelength of 630 nm.

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- The electromagnetic spectrum (EM spectrum) consists of electromagnetic radiation waves with a very wide range of wavelengths (and therefore, frequencies).
- EM radiation can propagate through a vacuum. It does not need a medium.
- EM radiation travels at the speed of light (c) in a vacuum; $c = 299~792~458~{
 m m.s}^{-1}$.
- EM radiation obeys the wave equation $v = f imes \lambda$ or $c = f imes \lambda$.
- EM radiation has a dual nature. It can behave both as a transverse wave and as a particle.
- EM radiation consists of an electric field and a magnetic field oscillating at right angles to each other and to the direction of wave propagation.

- EM waves with a higher frequency are more energetic.
- The discrete packages of energy transmitted by EM radiation are called photons.
- The energy of a photon is calculated using E = hf or $E = h \cdot \frac{c}{\lambda}$ where h is Planck's constant and

 $h=6.63 imes10^{-34}$ J.s·

Unit 2: Assessment

Suggested time to complete: 15 minutes

- 1. Calculate the wavelength of a radio wave with a frequency of 147.8 kHz?
- 2. What is the energy of a photon of EM radiation with a frequency of $3.75 imes 10^{13} \ {
 m Hz}$?
- 3. What is the energy of a photon of light with a wavelength of 450 nm?
- 4. What is the energy of a photon with a frequency of 12.56 GHz?
- 5. Two photons are detected. One has a wavelength of 276 nm and the other has a frequency of 400 THz. Which has the longer wavelength?

The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Exercise 2.1

1.
700 nm = 700 × 10⁻⁹ m

$$c = f \times \lambda$$

 $\therefore f = \frac{c}{\lambda}$
 $= \frac{300\ 000\ 000\ \text{m.s}^{-1}}{700 \times 10^{-9}\ \text{m}}$
 $= 428\ 571.43 \times 10^9\ \text{Hz}$
 $= 4.29 \times 10^{14}\ \text{Hz}$
2.
 $c = f \times \lambda$
 $\therefore \lambda = \frac{c}{f}$
 $= \frac{300\ 000\ 000\ \text{m.s}^{-1}}{2.1 \times 10^{18}\ \text{Hz}}$
 $= 1.43 \times 10^{-10}\ \text{m}$
3.
 $\therefore \lambda = \frac{c}{f}$
 $= \frac{300\ 000\ 000\ \text{m.s}^{-1}}{3.25 \times 10^{12}\ \text{Hz}}$
 $= 9.23 \times 10^{-5}\ \text{m}$

This wavelength falls outside of the visible spectrum of 700 nm to 400 nm and so cannot be detected by the human eye.

Back to Exercise 2.1

Exercise 2.2

1. E = hf $∴ E = 6.63 × 10^{-34} \text{ J.Hz}^{-1} × 4.67 × 10^{12} \text{ Hz}$ $= 3.096 × 10^{-21} \text{ J}$ 2. $630 \text{ nm} = 630 × 10^{-9} \text{ m}$ $E = h \cdot \frac{c}{\lambda}$ $∴ E = 6.63 × 10^{-34} \text{ J.Hz}^{-1} \cdot \frac{300\ 000\ 000\ \text{m.s}^{-1}}{630 × 1 - 9} \text{ m}$ $= 3.15 × 10^{-19} \text{ J}$

Back to Exercise 2.2

Unit 2: Assessment

```
1.
        147.8~\rm{kHz} = 1~478~000~\rm{Hz}
                         c=f	imes\lambda
                    \therefore \lambda = rac{c}{f}
                             = \frac{300\ 000\ 000\ {\rm m.s^{-1}}}{1\ 478\ 000\ {\rm Hz}}
                             = 202.98 \text{ m}
2.
            E = hf
        \therefore E = 6.63 	imes 10^{-34} 	ext{ J.Hz}^{-1} 	imes 3.75 	imes 10^{13} 	ext{ Hz}
                 = 2.48 	imes 10^{-20} \, \, {
m J}
3.
        450~{\rm nm} = 450 \times 10^{-9}~{\rm m}
                  E = h \cdot \frac{c}{\lambda}
              \therefore E = 6.63 	imes 10^{-34} \, \, {
m J.Hz^{-1}} \cdot {300 \,\, 000 \,\, 000 \,\, {
m m.s^{-1}} \over 450 	imes 10^{-9} \,\, {
m m}}
                        = 4.41 	imes 10^{-19} \, \, {
m J}
4.
        12.56~{\rm GHz} = 12.56 \times 10^9~{\rm Hz}
                         E = hf
                    \therefore E = 6.63 	imes 10^{-34} 	ext{ J.Hz}^{-1} 	imes 12.56 	imes 10^9 	ext{ Hz}
                              = 8.32 	imes 10^{-24} \, \, {
m J}
```

5.

400 THz = 400 × 10¹² Hz

$$c = f × \lambda$$

$$\therefore \lambda = \frac{c}{f}$$

$$= \frac{300\ 000\ 000\ \text{m.s}^{-1}}{400 \times 10^{12}\ \text{Hz}}$$

$$= 750 \times 10^{-9} \text{ m}$$

$$= 750 \text{ nm}$$

The photon with a frequency of $400~\mathrm{THz}$ has a longer wavelength.

Back to Unit 2: Assessment

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Unit 3: Applications of electromagnetic radiation

DYLAN BUSA



- Define lasers and give examples of their application.
- \cdot Explain and apply the photoelectric effect and give examples of its application.

What you should know

Before you start this unit, make sure you can:

- Explain what EM radiation is.
- Use the wave equation to calculate either the wavelength or frequency of EM radiation.
- Calculate the energy of a photon of EM radiation given either the wavelength or frequency. Refer to <u>unit 2</u> if you need help with any of this.

Introduction

In the previous unit we learnt what electromagnetic radiation is, how it forms and what properties it has. We noted that EM radiation covers an extremely wide range of wavelengths and frequencies and that we have classified different segments of the EM spectrum into different kinds of waves.

In this unit, we are going to explore the applications of these different kinds of EM radiation, but it is important to remember that a radio wave with a wavelength of a few hundred metres is fundamentally the same thing as visible light with a wavelength of 500 nm or a gamma ray with a wavelength of a few picometres. They are all EM radiation. They all obey the same laws. They are all formed in much the same way, and they propagate in the same way. All that is different is that they have different wavelengths and hence different frequencies and energies.

Applications of the EM spectrum

This section is based on <u>College Physics</u>. Authored by: OpenStax College. License: CC BY.¹

Figure 1 shows the full electromagnetic spectrum and which portions of it are used for specific applications. We can see, for example, that the part of the spectrum typically used for TV broadcasting also contains a

small section for FM radio and that AM radio uses radiation with a far longer wavelength (about 100 times longer).



Figure 1: The EM spectrum

Figure 1 also shows us that the different segments of the spectrum tend to overlap. There are no hard and firm rules about what is an X-ray and what is a gamma ray. It is all electromagnetic radiation with an almost infinite range of wavelengths. We only segment the spectrum like this so that we have some idea of the kind of wavelengths, frequencies and energies that are involved.

Table 1 lists the general segments of the EM spectrum and how waves in the range of each segment tend to be produced and what applications the EM radiation generally has. Gamma radiation, we can see, is generally produced when the unstable nuclei atoms lose energy by radiation. Hence another word for this is radioactive decay.

Type of EM radiation	Production	Applications	Life sciences aspect	Issues
Radio and TV	Accelerating charges	Communications, remote controls	MRI	Requires controls for band use
Microwaves	Accelerating charges and thermal agitation	Communications, ovens, radar	Deep heating	Cell phone use
Infrared	Thermal agitations and electronic transitions	Thermal imaging, heating	Absorbed by the atmosphere	Greenhouse effect
Visible light	Thermal agitations and electronic transitions	All pervasive	Photosynthesis, human vision	
Ultraviolet	Thermal agitations and electronic transitions	Sterilisation, cancer control	Vitamin D production	Ozone depletion, cancer causing
X-rays	Inner electronic transitions and fast collisions	Medical, security	Medical diagnosis, cancer therapy	Cancer causing
Gamma rays	Nuclear decay	Nuclear medicine, security	Medical diagnosis, cancer therapy	Cancer causing, radiation damage

Table 1: The production and applications of different types of EM radiation²

Let's take a closer look at each broad type of EM radiation and their applications.

2. Source: https://courses.lumenlearning.com/physics/chapter/ 24-3-the-electromagnetic-spectrum/ - CC-BY 4.0

Radio waves

Radio waves are the longest wavelength, lowest frequency and lowest energy EM radiation. They are produced on Earth by passing alternating currents through wires and other conductors. This part of the spectrum gets its name because of the major applications of radiation in this range – that of transmitting information, first by radio transmitters and receivers and later by television systems.

AM radio stations use frequencies between 540 kHz and 1 600 kHz. AM stands for **amplitude modulation**. A wave of a specific and constant frequency (the carrier wave) has its amplitude modified or modulated by an audio signal (see figure 2). The receiver is tuned to have a resonant frequency matching the carrier wave and uses the changes in amplitude to replicate the original audio.



Figure 2: Amplitude modulation

FM radio uses a slightly different technique. FM stands for frequency modulation. Instead of the carrier wave's amplitude being changed or modulated to carry the information, its frequency is changed (see figure 3).



Figure 3: Frequency modulation

FM uses frequencies between 88 MHz and 108 MHz. Because the frequency of the wave needs to be able to be changed by the range of human hearing (about 20 kHz) FM radio stations cannot be closer together on the dial than 20 kHz or 0.02 MHz. Interference (destructive and constructive) from other sources of EM radiation can more easily change a wave's amplitude than its frequency, therefore, FM radio suffers less from 'noise' than does AM radio. However, the lower frequencies used for AM mean that the longer wavelength waves are able to get around objects such as buildings and hills meaning that they tend to be better when broadcasting over longer ranges.

Traditional terrestrial television broadcasts typically use AM for the visuals and FM for the audio. TV uses the 54 MHz and 88 MHz and 174 MHz and 222 MHz bands. Satellite television and newer digital terrestrial broadcast use higher frequencies still (up to 1 000 MHz or more depending on regulations).

Mobile phone networks use frequencies in the range of $1.9~\mathrm{GHz}$ and magnetic resonance imaging (MRI) typically operates in the $100~\mathrm{MHz}$ range. It is a general rule of thumb that the higher the frequency of the radio wave, the more information it can carry.

Extremely low frequency (ELF) radio waves of about 1 _{kHz} are used to communicate with submarines because these waves are able to penetrate salt water quite well.

Many pieces of modern technology produce radio waves as a by-product. Perhaps the most pervasive are those produced by electricity transmission grids running at 50 Hz or 60 Hz. These have extremely long wavelengths of about 6 000 km but due to their low energy do not penetrate buildings and other structures well.



Figure 4: Transmission grids produce long wavelength EM radiation at $50~\mathrm{Hz}$ or $60~\mathrm{Hz}$

Radio waves are not only created on Earth by people. Certain astronomical phenomena also generate waves in the same range and these waves carry important information about these phenomena and the universe more generally. For example, the cosmic microwave background radiation has given astronomers one of the best pictures of how the universe probably formed and subsequently expanded (see figure 5).



Figure 5: The cosmic background radiation

But all the radio waves created on Earth can interfere with these astronomical efforts. This is why regulations control what wavelengths can be produced and where no EM radiation can be created at all. The two sites of the Square Kilometre Array (a huge radio telescope project) are being built in the middle of unpopulated deserts in Australia and South Africa (see figure 6) so that they are as far from any man-made EM radiation as possible.



Figure 6: The Square Kilometre Array (SKA) site, north of Carnarvon, Northern Cape, South Africa

Microwaves

Microwaves are the highest frequency EM radiation that can be produced, like radio waves, by electric currents in macroscopic circuits. They are so named because, next to radio waves, they have very small wavelengths. Typical frequencies range from 10^9 Hz to 10^{12} Hz.

The most known application of microwave radiation is the microwave oven for heating and cooking food. The energy in the EM radiation (2.45 GHz) is ultimately transferred to the water molecules in the food which gets hot and cooks.



45 GHz = 2.45 × 10° Hz

$$c = f \times \lambda$$

 $\therefore \lambda = \frac{c}{f}$
300 000 000 m

$$=rac{300\ 000\ 000\ \mathrm{m.s^{-1}}}{2.54 imes10^9\ \mathrm{Hz}} = 0.122\ \mathrm{m} = 122\ \mathrm{mm}$$

- 2. The diameter of the holes in the grid should be about 1 mm.
- 3. The holes are much smaller than the wavelength of the microwaves. Therefore, the grid effectively stops all the microwaves from escaping from the oven, reflecting them back inside the oven but the holes still allow you to see how your food is doing.

Cooking food is not all that microwaves are good for. Microwaves are also the basis of radar. Radar is a detection system that uses the reflection and doppler shift of EM radiation in the microwave range to determine the distance, angle, or speed of objects. It can be used to detect aircraft, ships, spacecraft, guided missiles, motor vehicles, weather formations, and terrain. Special radar systems are used to map the Earth and other planets. The smaller the wavelength used, the greater the detail that can be produced.

We noted above that the higher the frequency of the EM radiation, the more information it can carry. This is exploited in microwave-based communications, like satellite systems and internet backbones. But higher

frequencies mean lower wavelengths and so microwave communications require 'line of sight'. The receiver must be able to 'see' the transmitter without having any obstacles in the way.



Figure 7: A microwave communications tower

Infrared radiation

Infrared means 'below red' and includes those frequencies just below visible light and the upper end of the microwave range (although there is some overlap). When we think of infrared, we tend to think of night-vision scopes. These devices can detect the infrared radiation that is emitted by warm and hot objects.

Infrared radiation is produced by thermal motion – the vibration and rotation of atoms and molecules. The hotter something is, the more the atoms and molecules within it vibrate, the more infrared radiation it emits. The Sun, being really hot, emits a huge amount of infrared radiation. About half of the EM radiation received from the Sun is in the form of infrared radiation.

Half of this is absorbed by the Earth and then radiated back into space. Water molecules (H_2O) and carbon dioxide (CO_2) molecules are especially good at absorbing and re-emitting infrared radiation. Therefore, most of the radiation emitted from the Earth is absorbed by CO_2 and H_2O in the atmosphere and then radiated back to Earth or into outer space. This radiation back to Earth is known as the greenhouse effect but is generally a good thing. Without it, the surface temperature of the Earth would be about $40^{\circ}C$ lower. However, the more CO_2 and H_2O is in the atmosphere, the more radiation is absorbed and re-emitted back to Earth which seems to be resulting in higher global temperatures.

Visible light

About 44 of the EM radiation from the Sun is in the visible light range of 750 nm and 400 nm. Given the full spectrum that spans about 20 orders of magnitude, this is a very narrow band indeed. As figure 8 shows, we perceive light with the lowest frequency as red and light with the highest frequencies as violet. When the full spectrum is viewed together, we perceive white.



Figure 8: The visible light part of the EM spectrum

A common application of EM radiation in this part of the spectrum is Light Amplification by Stimulated Emission of Radiation (LASER). Not all lasers produce EM radiation in the visible spectrum, but many do (see figure 9).



Figure 9 Types of lasers with their operating wavelengths

Lasers are unique in that they produce EM radiation that is coherent. That means that the light produced is in a very narrow range of frequencies and intensities. Figure 10 shows the difference between several light sources. We can see that the light from a helium-neon laser (D) has a very narrow band of frequencies in it.



Figure 10: The very narrow spectrum of frequencies in a red light helium-neon laser (D) compared to other light sources

Lasers can be found everywhere – from CD players and pointers to, satellite communication systems, eye correction surgery, industrial cutting and distance detection.



Ultraviolet light

Ultraviolet (UV) means 'above violet' and extends from the highest frequency of visible light. The ultraviolet

spectrum includes waves from about 400 nm to 10 nm in wavelength. Some of the Sun's radiation is in the UV spectrum and is split into three categories:

- \cdot UV-A ($320-400\mathrm{nm}$) lowest energy
- UV-B (290 320nm)
- UV-C $(220 290 \mathrm{nm})$ highest energy.

Almost all UV-B and UV-C is absorbed by the ozone layer. Repeated and extensive exposure to any UV radiation can lead to skin cancer and other conditions such as cataracts. Exposure to UV radiation suffered by blast furnace workers in England in the late 1800s led many to become blind and the invention of the very first pair of sunglasses.

Staying in the shade, wearing protective clothing and wearing sunscreen are all effective ways of preventing over-exposure to UV radiation. However, some exposure is beneficial. UV-B, for example, stimulates the production of vitamin D, important for the prevention of some kinds of cancers and in general mood regulation.

UV radiation is typically used to sterilise objects and surfaces. In many clinics in South Africa, UV lamps are used to kill airborne pathogens especially those causing Tuberculosis (TB).



Figure 11: A UV lamp to kill airborne bacteria

X-rays

While the low-frequency end of the X-ray range overlaps with the ultraviolet, X-rays extend to much higher frequencies (and energies). They are created by very high voltage discharges into rarefied low pressure gasses. They were called X-rays because the scientists who discovered them did not know what they were.

Like UV radiation, X-rays have enough energy to damage living organisms. However, because they have higher energies, X-rays can penetrate deeper into the body than UV radiation. But used under controlled and targeted conditions, this same energy and penetration power can be used to kill cancer cells and 'see' inside the body. X-rays pass more easily through the soft tissues of the body than cartilage and bone. This difference is then processed to produce the X-ray image we are all familiar with (see figure 12).



Figure 12: A typical X-ray image

X-rays have wavelengths less than 0.01 nm. Therefore, they can also be used to detect the shapes of molecules based on how the radiation diffracts off the molecule. This kind of X-ray diffraction was most famously used by Francis Crick and James Watson in determining the shape of the DNA double-helix (see figure 13).


Gamma rays

Gamma rays are classified as any EM radiation emitted by a nucleus. They have extremely high frequencies and short wavelengths. Gamma rays are therefore even more penetrating and damaging to living organisms but can also be used in very controlled and targeted ways such as cancer therapies or for internal imaging. These kinds of applications fall under nuclear medicine.

Gamma rays occur naturally through a process of radioactive decay of certain materials in rocks and the soil and through many processes in space. Whether produced on Earth or in space, gamma rays are constantly bombarding our bodies.



The photoelectric effect

We learnt about the dual nature of light in the previous unit. Sometimes it behaves like a wave and sometimes it behaves like a stream of particles. For a long time, scientists thought that EM radiation was purely a continuous transverse wave (continuous flow of energy). It seemed to do all the things that other transverse waves do. This certainly helped to explain phenomena such as diffraction and interference.

However, in 1887 Heinrich Hertz (a German physicist and, yes, THAT Hertz) noticed that ultraviolet light, when shone on a metal plate, could cause sparks. The energy in the light gave the electrons in the metal enough energy to escape and the result was visible sparks.

The initial expectation was that, given enough time, EM radiation of any frequency would eventually cause sparks on any metal and that more sparks would result from a brighter light of the same frequency. It was thought that if the frequency or the intensity was initially too low, the electrons would gradually absorb more and more light energy until they were able to escape and cause sparks. However, experimental results showed this not to be the case.

Over time, it was noticed that different metals required light of different frequencies for the same effect. Then in 1902, Philipp Lenard (a Hungarian physicist) discovered that the maximum velocity with which electrons are ejected by ultraviolet light is entirely independent of the intensity of the light. How could this be? Surely the brighter the light, the more energy the electrons would be able to absorb and the greater their escape velocity would be.

A deep paradox existed at the heart of physics that no one could explain. That is until Albert Einstein took a break from his work on relativity. He proposed that people were thinking about the problem all wrong. He said that the EM radiation was not behaving like a continuous transverse wave but rather like a stream of tiny packets of energy called **quanta** (now called **photons**). All the photons in light of a particular frequency

had exactly the same amount of energy and it did not matter if there were only a few photons (dim light) or lots of photons (bright light).

He said that a photon gave up all its discrete energy to an electron in the metal and this is what allowed these electrons to escape and cause sparks (see figure 14).



Figure 14: The photoelectric effect

For many years, scientists were very sceptical of these ideas even though they explained the observed results very well. They were sceptical because Einstein solved the central paradox by introducing another one – that EM radiation can behave as a wave or as a particle, it just depends on how you look at it!

It turned out that Einstein was right and that his was the correct explanation of the **photoelectric effect**. It explained why:

- electrons are emitted immediately when EM radiation is shone on a metal,
- the intensity of EM radiation makes no difference to the maximum kinetic energy of the emitted electrons, and
- the photoelectric effect is only observed with different metals at different minimum frequencies.



- electromagnetic radiation shines on it.
 - $\cdot\;$ Electrons are emitted immediately when EM radiation is shone on a metal.
 - The intensity of EM radiation makes no difference to the maximum kinetic energy of the emitted electrons.
 - Different metals required different minimum frequencies for the photoelectric effect to be observed.

Note

If you have an internet connection, you should watch an excellent demonstration of the photoelectric effect called Photoelectric Effect Demonstration.

Photoelectric Effect Demonstration (Duration: 03.04)³



The photoelectric equation

The energy needed to eject an electron from a substance is called that substance's work function (W_0), and this is basically the amount of work that needs to be done to get an electron to escape. If the energy of a photon is less than the work function, then there is no emission and no sparks. It does not matter how many photons land on the metal, as one electron can only absorb one photon. Table 2 lists the W_0 values for a few common substances.

Element	Work Function($ m eV$)
Aluminium	4.08
Beryllium	5.0
Cadmium	4.07
Calcium	2.9
Carbon	4.81
Cesium	2.1
Cobalt	5.0
Copper	4.7
Gold	5.1
Iron	4.5
Lead	4.14
Magnesium	3.68
Mercury	4.5
Nickel	5.01
Niobium	4.3
Potassium	2.3, 2.29
Platinum	6.35
Selenium	5.11
Silver	4.26-4.73
Sodium	2.28, 2.36
Uranium	3.6
Zinc	4.3

Table 2: W_0 of some common substances

Note that the units in the table are given as electron volts (eV) where 1 $\,{
m eV} = 1.6 imes 10^{-19}\,$ J.

Remember, each photon has a discrete and fixed amount of energy depending on the radiation's frequency. These two quantities are related by Planck's constant in the equation E = hf or $(E = h \cdot \frac{c}{\lambda})$.

This is why there is a minimum frequency of EM radiation required for different substances with different work functions. We call this minimum frequency the cut-off (or threshold) frequency (f_0) . Therefore, we can say that $W_0 = hf_0$.



The cut-off frequency (f_0) is the minimum frequency of EM radiation needed to achieve the work function for a particular substance.

Now, we also know that energy is conserved. Any energy in a photon in addition to that required to achieve the work function goes into the electrons escape velocity or kinetic energy (E_K) . This excess photon energy is actually the **maximum** kinetic energy (E_{Kmax}) of an electron emitted right on the surface of the substance.

Electrons further down need additional energy to reach the surface before being ejected and so there is not the full excess amount available as kinetic energy.

We can therefore represent the total energy of a photon (*E*) that emits an electron for a given substance as: $E = W_0 + E_{Kmax}$ or $E_{Kmax} = hf - W_0$ (remember that E = hf)

This is the photoelectric equation.



We need to calculate the maximum kinetic energy i.e E_{Kmax} .

 $E_{Kmax} = hf - W_0$ but we have been given the wavelength not the frequency. Therefore, we can change the equation to $E_{Kmax} = h \cdot \frac{c}{\lambda} - W_0$.

$$egin{aligned} &W_0 = 6.9 imes 10^{-19} \,\, \mathrm{J} \ &\lambda = 245 \,\, \mathrm{nm} = 245 imes 10^{-9} \,\, \mathrm{m} \ &h = 6.63 imes 10^{-34} \,\, \mathrm{J.Hz^{-1}} \ &c = 300 \,\, 000 \,\, 000 \,\, \mathrm{m.s^{-1}} \ &E_{Kmax} = h \cdot rac{c}{\lambda} - W_0 \ &\therefore E_{Kmax} = 6.63 imes 10^{-34} \,\, \mathrm{J.Hz^{-1}} \cdot rac{300 \,\, 000 \,\, 000 \,\, \mathrm{m.s^{-1}}}{245 imes 10^{-9} \,\, \mathrm{m}} - 6.9 imes 10^{-19} \,\, \mathrm{J} \ &= 1.22 imes 10^{-19} \,\, \mathrm{J} \end{aligned}$$

Therefore, the maximum kinetic energy that any emitted electron can have will be $1.22x10^{-19}$ J.



UV light ($f = 4.6x10^{15}$ Hz) is incident to a sodium surface ($W_0 = 4.41 \times 10^{-19}$ J). Would any surface electrons be emitted?

Solution

 (\bigcirc)

We need to determine if the energy in a photon is greater than the work function.

$$egin{aligned} f &= 4.6 imes 10^{15} \,\, {
m Hz} \ h &= 6.63 imes 10^{-34} \,\, {
m J.Hz^{-1}} \ E &= hf \ dots &= 6.63 imes 10^{-34} \,\, {
m J.Hz^{-1}} imes 4.6 imes 10^{15} \,\, {
m Hz} \ &= 3.05 imes 10^{-18} \,\, {
m J} \end{aligned}$$

Therefore, electrons will be emitted as the energy of each photon is greater than the work function of sodium.



Exercise 3.1

Question 1 adapted from Everything Science Grade 12 Worked Example 3: Photoelectric effect

- 1. A metal surface is illuminated with ultraviolet light of wavelength 330 nm. Electrons are emitted from the metal surface. The minimum amount of energy required to emit an electron from the surface of this metal is $3.5x10^{-19}$ J.
 - a. Name the phenomenon illustrated above.
 - b. What is the term for 'the minimum amount of energy required to emit an electron from the surface'?
 - c. Calculate the frequency of the ultraviolet light.
 - d. Calculate the kinetic energy of a photoelectron emitted from the surface of the metal when the ultraviolet light shines on it.
 - e. The intensity of the ultraviolet light illuminating the metal is now increased. What effect will

this change have on the kinetic energy of the emitted photoelectrons? (Write down only INCREASES, DECREASES or REMAINS THE SAME.)

- f. The intensity of the ultraviolet light illuminating the metal is now increased. What effect will this change have on the number of photoelectrons emitted per second? (Write down only INCREASES, DECREASES or REMAINS THE SAME.)
- 2. A light of an unknown wavelength is incident to silver foil. The light has only enough energy to eject electrons from the silver foil but not enough to give them kinetic energy.
 - a. If the same light is incident on copper foil, would electrons be ejected? ($W_0 = 7.5 x 10^{-19}$ J)
 - b. If the same light is incident on silicon, would electrons be ejected? ($W_0 = 1.8x10^{-19}$ J)
 - c. If the intensity of the light on the silver foil is increased, what happens?
 - d. If the frequency of the light incident on the silver foil is increased, what happens?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Every part of the EM spectrum has practical applications.
- · Laser stands for Light Amplification by Stimulated Emission of Radiation.
- Laser light is unique in that it is coherent all the light is in phase and of a very similar wavelength and frequency.
- The photoelectric effect was key to helping scientists understand the dual nature of EM radiation that it can behave like a continuous transverse wave and like a stream of particles called photons.
- The photoelectric effect occurs when the energy in photons of EM radiation is sufficient to knock electrons out of a substance.
- The work function (W_0) is the minimum amount of energy required for a substance to emit an electron.
- The cut-off frequency or threshold frequency (f_0) is the minimum frequency of EM radiation required for the photoelectric effect to occur.
- The photoelectric equation is $E = W_0 + E_{Kmax}$ or $E_{Kmax} = hf W_0$.

Unit 3: Assessment

Suggested time to complete: 25 minutes

Question 1 adapted from NC(V) Physical Sciences Paper 1 November 2019 question 11

- 1. Ultraviolet light with a frequency of $3x10^{15}$ Hz strikes a metal surface and ejects photoelectrons that have a maximum kinetic energy of $9.76x10^{-19}$ J.
 - a. Calculate the minimum amount of energy required to emit an electron from the metal.
 - b. Calculate the threshold frequency of the metal.
 - c. What effect will an increase in frequency have on the: (only write INCREASES, DECREASES or NO EFFECT)
 - i. speed of the photoelectrons?

- ii. number of photoelectrons released?
- d. Explain how the intensity of the incident light affects the number of photoelectrons ejected from the metal surface.

Question 2 adapted from NC(V) Physical Sciences Paper 1 February 2019 question 11

- 2. Ultraviolet light with a wavelength of 280 nm is shone on a metal surface. Electrons are emitted from the metal surface. The minimum amount of energy required to emit electrons from the surface of the metal is $3.5x10^{-19}$ J.
 - a. What is the NAME given to the minimum amount of energy required to emit electrons from the surface of the metal?
 - b. Calculate the frequency of the ultraviolet light.
 - c. Calculate the kinetic energy of the electrons emitted from the surface of the metal.
 - d. The intensity of the ultraviolet light illuminating the metal is decreased. What effect will this have on the following: (write down only INCREASES/ DECREASES/REMAINS THE SAME)
 - i. the kinetic energy of the emitted electrons?
 - ii. the number of electrons emitted per second?

The <u>full solutions</u> are at the end of the unit.

Unit 3: Solutions

Exercise 3.1

1.

- a. This is the photoelectric effect.
- b. The work function.

c.

```
\begin{split} \lambda &= 330 \text{ nm} = 330 \times 10^{-19} \text{ m} \\ c &= 300 \ 000 \ 000 \ \text{m.s}^{-1} \\ c &= f \times \lambda \\ \therefore f &= \frac{c}{\lambda} \\ &= \frac{300 \ 000 \ 000 \ \text{m.s}^{-1}}{330 \times 10^{-19} \ \text{m}} \\ &= 9.09 \times 10^{20} \ \text{Hz} \\ \text{d.} \\ \end{split}
\begin{aligned} W_0 &= 3.5 \times 10^{-19} \ \text{J} \\ f &= 9.09 \times 10^{20} \ \text{Hz} \\ h &= 6.63 \times 10^{-34} \ \text{J.Hz}^{-1} \\ E_{Kmax} &= hf - W_0 \\ &= 6.63 \times 10^{-34} \ \text{J.Hz}^{-1} \times 9.09 \times 10^{20} \ \text{Hz} - 3.5 \times 10^{-19} \ \text{J} \\ &= 6.03 \times 10^{14} \ \text{J} \end{aligned}
e. REMAINS THE SAME
f. INCREASES
```

- a. No. The work function of copper is greater than silver. Therefore, the photons would need to have more energy to eject electrons from copper.
- b. Yes. The work function of silicon is less than silver. Therefore, the photons would have enough energy to eject electrons from silicon.
- c. More electrons per second are emitted but they have the same maximum kinetic energy.
- d. The same number of electrons would be emitted per second but each of these would have greater maximum kinetic energy.

Back to Exercise 3.1

Unit 3: Assessment

```
1.
       а.
                  f=3	imes 10^{15}~{
m Hz}
                  h = 6.63 	imes 10^{-34} \, \, {
m J.Hz^{-1}}
                  E = hf
              \therefore E = 6.63 \times 10^{-34} \text{ J.Hz}^{-1} \times 3 \times 10^{15} \text{ Hz}
                     = 1.989 	imes 10^{-18} \,\, {
m J}
              E_{Kmax} = 9.76 	imes 10^{-19} \,\, {
m J}
                      E = W_0 + E_{Kmax}
                \therefore W_0 = E - E_{Kmax}
                           = 1.989 \times 10^{-18} \ J - 9.76 \times 10^{-19} \ J
                           = 10.13 \times 10^{-19} \, \rm ~J
      b.
                 E = 10.13 	imes 10^{-19} \,\, {
m J}
                 h = 6.63 	imes 10^{-34} \, \, {
m J.Hz^{-1}}
                E = hf
              \therefore f = \frac{E}{h}
                    =rac{10.13	imes 10^{-19}\,\,{
m J}}{6.63	imes 10^{-34}\,\,{
m J.Hz}^{-1}}
                     =1.53	imes10^{15}~{
m Hz}
```

c.

i. Increase

- ii. No effect
- d. The greater the intensity, the more photons hit the material per second. The more photons that hit the material, the more electrons will be emitted by the material.

2.

a. Work function

b.

$$\begin{split} \lambda &= 280 \text{ nm} = 280 \times 10^{-9} \text{ m} \\ c &= 3 \times 10^8 \text{ m.s}^{-1} \\ c &= f \times \lambda \\ \therefore f = \frac{c}{\lambda} \\ &= \frac{3 \times 10^8 \text{ m.s}^{-1}}{280 \times 10^{-9} \text{ m}} \\ &= 1.07 \times 10^{15} \text{ Hz} \end{split}$$
c.
$$h &= 6.63 \times 10^{-34} \text{ J} \\ W_0 &= 3.5 \times 10^{-19} \text{ J} \\ f &= 1.07 \times 10^{15} \text{ Hz} \\ E_{Kmax} &= hf - W_0 \\ &= 6.63 \times 10^{-34} \text{ J} \times 1.07 \times 10^{15} \text{ Hz} - 3.5 \times 10^{-19} \text{ J} \\ &= 3.59 \times 10^{-19} \text{ J} \end{split}$$

d.

- i. Remains the same
- ii. Decreases

Back to Unit 3: Assessment

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SUBJECT OUTCOME VI ELECTRICITY AND ELECTRONICS: STATE, EXPLAIN AND APPLY PRINCIPLES USED IN ELECTRODYNAMICS

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Subject outcome

Subject outcome 4.1: State, explain and apply principles used in electrodynamics



Learning outcomes

- Identify and differentiate between stationary and moving electrical machines (transformer, generator, and motor)
- Describe the efficiency of stationary and moving electrical machines
- \cdot Define single-phase and three-phase and identify its applications.
- \cdot Define and calculate energy transfer in an electrical circuit.
- \cdot Define and calculate electrical power and identify its application on tools, etc.
- Determine the relationship between current, pd, resistance and power.



Unit 1 outcomes

By the end of this unit you will be able to:

- Identify and differentiate between stationary and moving electrical machines (transformer generator and motor).
- \cdot $\,$ Describe the efficiency of stationary and moving electrical machines.
- Define single-phase and three-phase.
- · Identify the applications of single- and three-phase electricity.



Unit 2 outcomes

By the end of this unit you will be able to:

• Define and calculate energy transfer in an electrical circuit.

- Define and calculate electrical power.
- Identify the application of power on tools, etc.
- $\cdot\;$ Determine the relationship between current, pd, resistance and power.

Unit 1: Generating electricity

LEIGH KLEYNHANS



Unit outcomes

By the end of this unit you will be able to:

- · Identify and differentiate between stationary and moving electrical machines (transformer, generator and motor).
- Describe the efficiency of stationary and moving electrical machines.
- Define single-phase and three-phase.
- · Identify the applications of single- and three-phase electricity.

What you should know

Before you start this unit, make sure you can:

- Describe and apply electromagnetic induction to produce current by using a changing magnetic field. Refer to level 3 subject outcome 4.2 unit 2 if you need help with this.
- Describe how electromagnetic induction is applied in dynamos and transformers. Refer to level 3 subject outcome 4.2 unit 2 if you need help with this.

Introduction

An electrical machine transforms energy. This can be either electrical energy converted into mechanical energy which happens in a motor or the conversion of mechanical energy into electrical energy which happens in a generator. A transformer is also an electrical machine that converts voltage and current to different levels. You have already learnt how generators and transformers apply the principle of electromagnetic induction in their operation. In this unit we will classify electrical machines according to whether they involve moving parts or not, as well as introduce the principle of single- and three-phase electricity.

Types of electrical machines

Electrical machines are classified into two main types:

- stationary electrical machines
- dynamic (moving) electrical machines.

Stationary electrical machines

A stationary electrical machine is a machine which does not have any moving parts. The components remain stationary throughout its operation.

A transformer is an example of a stationary electrical machine. It is a machine because there is a conversion of energy between the windings (turns) of the transformer. Electrical energy is converted into magnetism in one coil of the transformer and then the magnetism is converted back into electrical energy in the other coil. The ratio of the number of turns in the primary coil to the number of turns in the secondary coil determines how this process increases or decreases the AC voltage and current levels.

The primary and secondary coils are both wound around a stationary iron core. There are no moving parts involved in the energy transformation. It is the varying AC current in the primary winding, which creates a varying magnetic flux in the iron core of the transformer. This varying magnetic flux induces an emf in the secondary winding of the transformer, resulting in an AC current at the output.



Figure 1: A typical construction of a simple transformer with two coils wound on a ferromagnetic core; there are no moving parts

Many cell phones, laptops, video games, power tools, and small appliances have a transformer built into their plug-in unit that changes 120 V or 240 V into the voltage required by the device.

Dynamic (moving) electrical machines

Dynamic electrical machines consist of moving parts as well as stationary parts. There are two types of dynamic electrical machines: generators and motors .

Generators are electrical machines which convert mechanical energy into electrical energy. You have learnt about Faraday's law, which states that a conductor placed inside a varying magnetic field will experience an induced emf. In other words, moving a conductor in a static magnetic field will induce emf in the conductor and current electricity will be produced.



Figure 2: A simple generator

When the coil in the generator in figure 2 is rotated, the magnetic flux changes from a maximum to zero, inducing an emf, which drives a current through an external circuit. The moving part is the coil that is rotated inside the magnetic field. The source of mechanical energy that turns the coil can be falling water (hydropower), rising steam produced by the burning of fossil fuels, or the kinetic energy of wind. The resultant induced current (electrical energy) flows through the conductor.

A **motor** is a dynamic electrical machine which converts electrical energy into mechanical energy – the exact opposite of a generator. Electrical motors have a moving part called the rotor and a stationary part called the stator.

Electrical motors generate a mechanical force as a result of the interaction between a magnetic field and current in a conductor placed in the magnetic field. Motors consist of loops of wire (the rotor) in a magnetic

field (the stator). When current is passed through the loops, the magnetic field exerts a torque on the loops, which rotates a shaft. Electrical energy is converted to mechanical energy in the process.



Figure 3: Torque (au) on a current loop in a motor

This effect can be seen in toy cars where the rotating shaft is connected to the wheels of the car and a battery supplies the current. Motors are also used in washing machines, tumble driers and fans to create the rotating movement in these appliances.



Exercise 1.1

- 1. What happens in an electric motor?
 - A. An electric motor transforms electrical energy into mechanical energy.
 - B. An electric motor transforms mechanical energy into electrical energy.
 - C. An electric motor transforms chemical energy into mechanical energy.
 - D. An electric motor transforms mechanical energy into chemical energy.
- 2. What happens in a transformer?
 - A. A transformer transforms electrical energy into mechanical energy.
 - B. A transformer transforms magnetism into electrical energy.
 - C. A transformer transforms chemical energy into mechanical energy.
 - D. A transformer transforms mechanical energy into chemical energy.
- 3. What happens in a generator?
 - A. A generator transforms electrical energy into mechanical energy.
 - B. A generator transforms mechanical energy into electrical energy.
 - C. A generator transforms chemical energy into mechanical energy.
 - D. A generator transforms mechanical energy into chemical energy.

The <u>full solutions</u> are at the end of the unit.

Efficiency of electrical machines

Efficiency is a measurement of how well a machine carries out the transformation of energy from one kind into another. It is the ratio of the output energy to the input energy, expressed as a percentage.

 $\% Efficiency = \frac{output \ energy}{input \ energy} \ \mathbf{x} \ 100$

It is simply not possible to make a machine 100% efficient. There will always be some wasted energy. This means conversion into types of energy other than the type of energy required from the machine.

In dynamic electrical machines (motors and generators) the moving parts will experience friction resulting in some conversion of energy into heat. Friction results from surfaces rubbing together. Friction can be minimised by using bearings or lubricants such as grease, oil, or fine powder. This will reduce the amount of wasted energy.

Efficiency in a dynamic electrical machine can also be reduced by what is referred to as windage. Windage results from the rotating parts driving air around a motor, inside or out. Fans inside a motor increase the windage but help to dissipate heat, so it has an advantage as well as a disadvantage.

In stationary electrical machines (transformers), there is no loss of energy as a result of friction because there are no moving parts. An ideal transformer would therefore be 100% efficient. However, in practice, energy is dissipated due to resistance in the wires of the windings and to magnetic effects in the core. Transformers are, in general, highly efficient, and large power transformers in the national grid may attain efficiencies as high as 99.75%. Small transformers such as those that power consumer electronics may be about 85% efficient. The small percentage of dissipated energy is in the form of heat. You may have experienced this when you touch a transformer after a device has been operating for some time, it will feel warm.

Regardless of which type of electric machine is chosen for a particular job, efficiency is important for several reasons. Firstly, and perhaps most obvious, a less efficient machine will cost more to operate. Secondly, the energy 'losses' in the machine are converted to heat, which raises the operating temperature of the machine. The life of insulation is strongly related to the operating temperature. Thus, the machine must be designed to tolerate the heat created.



Exercise 1.2

- 1. How is efficiency of an electrical machine calculated?
- 2. Complete the following table:

Type of electrical machine	Primary type of output energy	Type of wasted energy
motor		
generator		
transformer		

The <u>full solutions</u> are at the end of the unit.

Single and three-phase electricity supply

Single phase electricity supply refers to when there is one conducting wire in the AC current circuit. This is the kind of electricity supply from a generator as illustrated in figure 2 and the operation of which you have studied in a previous unit. As the coil is rotated through the magnetic field, an emf, and consequently

a current, is induced. Because the change in the magnetic flux is not constant, the induced emf and current are not constant, but fluctuate at a regular frequency. This can be represented in a graph as illustrated in figure 4.



Figure 4: A graph to illustrate the changing induced emf (voltage) in a generator

The disadvantage of single-phase electricity, as seen in the graph in figure 4, is that twice every cycle the voltage will be zero. This problem can be solved by using a generator with three coils, each at 120° to each other and all rotating through the magnetic field.



Figure 5: A generator producing three-phase electricity

As a result of this set-up, the electric current is carried in three separate conducting wires; P1, P2 and P3, referred to as three-phase. At no time will there be zero voltage, and the supply is relatively constant.



Figure 6: A graph of three-phase electricity

The graph shows the three separate wave forms for each conducting wire.

Residential homes are usually served by single-phase power supply, while commercial and industrial facilities usually use three-phase supply. The key difference between single-phase and three-phase is that three-phase better accommodates higher loads. Single-phase is sufficient for lighting and heating, however, three-phase would be required to run large electric motors. Three phase supply is also more efficient than single-phase. Single phase AC has the advantage that it only requires two wires.



- 2. Define three-phase electricity supply.
- 3. Give an example of where single-phase and where three-phase electricity supply would be used.

The <u>full solutions</u> are at the end of the unit.

Summary

- A stationary electrical machine is a machine which does not have any moving parts. The components remain stationary throughout its operation. An example is a transformer.
- Dynamic electrical machines consist of moving parts as well as stationary parts. Examples are motors and generators.
- Efficiency is a measurement of how well a machine carries out the transformation of energy from one kind into another. It is calculated using the formula:

 $\% Efficiency = \frac{output \ energy}{input \ energy} \ \mathbf{x} \ 100$

- No electrical machines will be 100% efficient because of the conversion of some electrical energy into heat energy during operation.
- Moving electrical machines will be less efficient than stationary electrical machines because of friction between moving parts.
- Single-phase electricity supply uses a single conducting wire, the voltage and current fluctuate and can be represented in a single wave form.
- Three-phase electricity supply uses three conducting wires to supply current in order to keep the current consistent. It is represented in a triple wave form.
- Single-phase electricity is used when the demand of the load is low and three-phase electricity is used in high demand situations.

Unit 1: Assessment

Suggested time to complete: 15 minutes

- 1. What makes stationary electrical machines different from moving electrical machines?
- 2. Why are transformers classified as stationary electrical machines?
- 3. State three sources of mechanical energy used to rotate the moving parts of a generator.
- 4. Name the component of a motor that:

- a. is stationary
- b. is moving.
- 5. Why are electrical machines not 100% efficient?
- 6. Why would it be important to know the efficiency of an electrical machine?
- 7. Tabulate the features of single-phase and three-phase electricity supply under the following headings: operation, wave form, construction, efficiency, cost, and applications.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

- 1. A
- 2. B
- 3. B

Back to Exercise 1.1

Exercise 1.2

1. The ratio of the output energy to the input energy converted to a percentage; use the formula: $\% Efficiency = \frac{output \ energy}{input \ energy} \ge 100$

2.

Type of electrical machine	Primary type of output energy	Type of wasted energy
motor	kinetic/mechanical	heat
generator	electrical	heat
transformer	electrical	heat

Back to Exercise 1.2

Exercise 1.3

- 1. Single-phase electricity supply uses a single conducting wire, the voltage and current fluctuate and can be represented in a single wave form.
- 2. Three-phase electricity supply uses three conducting wires to supply current in order to keep the current consistent. It is represented in a triple wave form.
- 3. Single-phase: homes, shops, offices for heating and lighting Three-phase: factories, industrial plants using heaving machinery

Back to Exercise 1.3

Unit 1: Assessment

- 1. There are no moving parts in a stationary electrical machine, whereas a moving electrical machine has both moving and static parts.
- 2. Transformers have no moving parts.
- 3. falling water, rising steam, wind
- 4.
- a. stator
- b. rotor
- 5. Not all the input energy is converted into output energy. Some energy is used to overcome friction in the moving parts and is transformed into heat.
- 6. The efficiency gives an indication of how much energy is converted into useful energy and how much will be wasted during operation of the machine.
- 7.

Feature	Single-phase	Three-phase
Operation	Uses a single conductor	Uses three conductors
Wave cycle	Only one wave cycle	Three wave cycles
Construction	Simple	Complex
Efficiency	Lower	Higher
Cost	Lower	Higher
Applications	Low loads – houses, offices	Very high loads – machinery

Back to Unit 1: Assessment

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Unit 2: Energy transfer in electrical circuits

LEIGH KLEYNHANS



What you should know

Before you start this unit, make sure you can:

- Identify components in electrical circuits. Refer to <u>level 3 subject outcome 4.3 unit 2</u> if you need help with this.
- Identify the relationship between current, voltage and resistance in electrical circuits. Refer to <u>level 3</u> <u>subject outcome 4.3 unit 1</u> if you need help with this.

Introduction

In this unit you will be learning about energy transfer in electrical circuits as well as applying the principle of power to electrical circuits.

Energy transfer in electrical circuits

A source of energy is required to drive current round a complete circuit. This is provided by batteries or an electrical generator. The energy is used to do work on the electrons in the circuit. The emf of the battery or the induced emf in a generator is the total energy supplied per coulomb of charge. It is also referred to as potential difference as the electrical potential energy of the charge changes as it moves through the battery or coils of the generator.

This can be written as a formula: $V = \frac{W}{q}$

Where:

V is the voltage (emf) in volts W is the work (energy) in joules (J) q is the charge in coulombs (C)

For example:

- A 4.5 V cell supplies 4.5 J of electrical energy to every coulomb of charge that passes through the cell.
- A 9 V cell supplies 9 J of electrical potential energy to every coulomb of charge that passes through the cell.



- \cdot Calculate the emf of a battery if it supplies 250 J of energy when 10 C of charge pass through it.
- \cdot Calculate how much energy is transferred when 0.5 C of charge passes through a cell of 6 V.

The <u>full solutions</u> are at the end of the unit.

Power in electrical circuits

As charges pass through resistors in an electrical circuit the electrical energy they carry is transformed into other types of energy in the resistors. Power is a measure of how rapidly this energy transformation or work is done, and is defined as the rate at which the work is done, or work done per unit time. Work is measured in joules (J) and time in seconds (s) so power will be J.s which we call a watt (W). The following formula can be used to calculate power:

$$P = \frac{W}{\Delta t}$$

Where:

W is the work/energy in joules (J) Δt is the time in seconds (s) P is the power in watts (W)

Application of power rating

Electrical appliances have a power rating which is an indication of how fast they can convert electrical energy into other types of energy. A heater that has a power rating of 850 W means that 850 J of electrical energy is converted into heat energy every second.

When purchasing an electrical appliance, it is important to choose the correct power rating for the purpose. For example, if you need to heat a large room, it would be better to buy a heater with a power rating of $2\ 000\ W$ rather than one with a power rating of $1\ 000\ W$ as the rate at which electrical energy is transformed into heat energy is double in the heater with the higher power rating. However, you need to bear in mind that the electricity consumption and therefore the cost of running the higher rated heater will be double that of the one with the lower power rating.

As another example, a battery-operated drill may have a power rating of 400 W whereas one using mains supply may have a power rating of 1 000 W. If you are drilling holes in soft wood, the battery-operated drill may supply sufficient mechanical energy per second, however, if you are drilling into concrete, the energy provided per second would not be sufficient and you would be better advised to use the higher rated drill.

Using multiple appliances with high power ratings at the same time may overload circuits and cause the electricity supply to trip.

$$\begin{aligned}
 For the equation of the equation$$



Exercise 2.2

- 1. How long will it take a kettle to produce $3 \ge 10^6$ J of energy if it has a power rating of 1 500 W?
- 2. How much mechanical energy is produced by a fan with a power rating of $400~{
 m W}$ in one hour?
- 3. Two light bulbs, one rated at 200 W and the other at 60 W are available to light a dark parking area. Explain which light bulb should be chosen.

The <u>full solutions</u> are at the end of the unit.

The relationship between current, voltage (PD), resistance and power

The energy used by a component in a circuit with a particular resistance (R) depends on the potential difference (V) across the component, the current (I) through it and the length of time (Δt) for which the current flows. The relationship between these variables is given in the following formulae:

$$W = V I \Delta$$
 $W = I^2 R \Delta t$ $W rac{V^2}{R} \Delta t$

The relationship of these variables with the power of the component can be determined by combining these formulae with the defining formula for power:

$$P = rac{W}{\Delta t} ext{ or } W = P \Delta t$$

This gives three formulae that can be used to calculate the power of an electrical component:

$$P = VI$$
 $P = I^2 R$ $P = rac{V^2}{R}$

All these formulae are very useful when determining the relationship between the variables in an electrical appliance.

These relationships are:

- From P = VI, the power is directly proportional to the voltage and the current.
- From $P = I^2 R$, the power is directly proportional to the square of the current and the resistance.
- From $P = \frac{V^2}{R}$, power is directly proportional to voltage and inversely proportional to resistance.

Example 2.3

The resistance of a resistor is 27 Ω and the current is 3.3 A. How much energy is dissipated in the resistor in 35 s?

Solution

Step 1: List the given information and check the units

I = 3.3 A $R = 27 \Omega$ $\Delta t = 35 \text{ s}$ W = ?Step 2: Choose the correct formula

 $W = I^2 R \Delta t$

Step 3: Substitute the values and calculate the answer

 $W = (3.3)^2 (27)(35)$ = 10 291.05 J

Example 2.4

A kettle connected to a 220 V power supply and a resistance of 50 Ω takes three minutes to boil water. How much energy is used?

Solution

 \mathbb{Q}

Step 1: List the given information and check the units

$$V=220 ext{ V}$$

 $R=100 ext{ \Omega}$
 $\Delta t=3 ext{ minutes}=180 ext{ s}$
 $W=?$

Step 2: Choose the correct formula

$$W = rac{V^2}{R} \Delta t$$

Step 3: Substitute the values and calculate the answer

$$W = rac{220^2}{50}(180) = 174\ 240\ {
m J}$$

Example 2.5

Given that a circuit component has a voltage of 5 V and a resistance of 2 Ω , calculate the power of the component.

Solution

Q

Step 1: List the given information and check the units

$$V = 5 V$$

 $R = 2 \Omega$
 $P = ?$

Step 2: Choose the correct formula

$$P = \frac{V^2}{R}$$

Step 3: Substitute the values and calculate the answer

$$P = \frac{5^2}{2}$$
$$= 12.5 \text{ W}$$



Example 2.6

A resistor has a resistance of 15 Ω and the current through it is 400 mA. What is the power of the resistor?

Solution

Step 1: List the given information and check the units

I = 400 mA = 0.4 A $R = 15 \Omega$ P = ?

Step 2: Choose the correct formula

$$P = I^2 R$$

Step 3: Substitute the values and calculate the answer

 $P = (0.4^2)(15)$ = 2.4 W

Summary

- Batteries and generators supply energy to electrical circuits.
- The emf of a battery or generator is defined as the total energy supplied per coulomb of charge. It can be calculated using the formula: $V = \frac{W}{q}$.
- Power is the rate at which energy is transferred or transformed. It can be calculated using the formula: $P = \frac{W}{t}.$
- Power is measured in the units watts (W) which are equivalent to $J_{.s}^{-1}$.
- The energy transformed by a component in an electrical circuit is determined by the voltage across it, the current flowing through it and its resistance. The following formulae can be used to calculate this energy transformation:

$$W = VI\Delta t$$
 $W = I^2 R\Delta t$ $W = \frac{V^2}{R}\Delta t$

• The following formulae show the relationship between the power, the voltage, the current and the resistance of an electrical device:

$$P = VI$$
 $P = I^2 R$ $P = rac{V^2}{R}$

Unit 2: Assessment

Suggested time to complete: 30 minutes

- 1. Energy of $4 \ge 10^3$ J is used to move $2 \ge 10^5$ C of charge from one point in a circuit to another point. What is the potential difference between the two points?
- 2. A battery has a potential difference of 6 V. When it is connected into a circuit, 3 C of charge flows

through a conductor. Calculate the energy transferred to the conductor.

- 3. A current of 0.5 A flows through a conductor for 5 s. During that time, 5 J of energy are transferred to the conductor. Calculate the potential difference across the conductor.
- 4. A resistor of 5 Ω has a current of 0.2 A passing through it for five minutes. Calculate the energy transferred to the resistor, by the current, during that time.
- 5. A microwave has a power rating of 800 W.
 - a. What does this mean?
 - b. If the microwave is connected to a 220 V power source, calculate its resistance.
- 6. A light bulb is rated 60 W: 220 V
 - a. What does this mean?
 - b. What current passes through the bulb when it is operating under correct conditions?
 - c. How much energy does the bulb transfer in 30 s?
- 7. What is the power of a $1 \ge 10^2$ MV lightning bolt having a current of $2 \ge 10^4$ A?
- 8. How much work is done in the starter motor of a large truck that draws 250 A of current from a 24 V battery in 0.5 s?
- 9. A charge of 4 C passes through a pocket calculator's solar cells in four hours. What is the power output, given the calculator's voltage output is 3 V?
- 10. Find the power dissipated in each of these extension cords through which 5 A passes:
 - a. an extension cord having a $0.06~\Omega$ resistance
 - b. a cheaper cord with a resistance of $0.03~\Omega$.
- 11. A circuit transfers 3000 J energy to a heater. If the current through the heater is 15 A and the potential difference across it is 200 V calculate the time required for the energy to be transferred.
- 12. The power dissipated in a resistor is given by $P = \frac{V^2}{R}$, which means power decreases if resistance

increases. Yet power is also given by $P = I^2 R$, which means power increases if resistance increases. Explain why there is no contradiction here.

The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Exercise 2.1

1.

$$v = \frac{W}{q}$$

$$= \frac{250}{10}$$

$$= 25 \text{ V}$$
2.

$$W = Vq$$

$$= 6 \text{ x } 0.5$$

$$= 3 \text{ J}$$

Back to Exercise 2.1

Exercise 2.2

1.

$$\Delta t = \frac{W}{P}$$
$$= \frac{3 \times 10^6}{1 500}$$
$$= 2 000 \text{ s}$$

2.

 $W = P \ge \Delta t$

= 400 x 3 600 (hours must be converted to seconds)

 $= \ 1 \ 440 \ 000 J$

 \mathbf{s}

3. The 200 W light bulb would be better as it converts 200 J of electrical energy into light energy per second so the light would be brighter. The $60~{
m W}$ bulb would only convert $60~{
m J}$ of energy per second.

Back to Exercise 2.2

Unit 2: Assessment

1. $V = \frac{W}{W}$ $= \frac{4 \, \mathrm{x} \, 10^3}{2 \, \mathrm{x} \, 10^5}$ $= 0.02 \ V$ 2. W = Vq $= 6 \ge 3$

3.

$$W = VI\Delta t$$
$$V = \frac{W}{I\Delta t}$$
$$= \frac{5}{0.5 \text{ x 5}}$$
$$= 2 \text{ V}$$

= 18 J

4.

 $W=I^2R\Delta t$ $= (0.2)^2(5)(3000)$ (convert minutes to seconds) $= 60 \mathrm{J}$

5.

a. 800 J of energy is transformed per second

b.

$$P = \frac{V^2}{R}$$
$$R = \frac{V^2}{P}$$
$$= \frac{220^2}{800}$$
$$= 60.5 \ \Omega$$

6.

a. When connected to a 220 V power supply, 60 J of energy will be transformed per second.

```
b.
            P = VI
            I = \frac{P}{V}
              =\frac{60}{220}
              = 0.27 A
      c.
            W = P\Delta t
                = 60 \ge 30
                = 1 800 J
 7.
      P = VI
         = (1 \ge 10^5)(2 \ge 10^4) (convert MV to V
         = 2 \times 10^9 W
 8.
      W = VI\Delta t
          = (24)(250)(0.5)
         = 3 000 J
 9.
      V=rac{W}{q}
      W = Vq
         = 3 x 4
          = 12 \text{ J}
       \mathrm{P}=rac{W}{\Delta t}
                   12
          =rac{12}{(4 \ge 60 \ge 60)}
          = 8.33 \text{ x } 10^{-4} \text{ W}
10.
       a.
            P = I^2 R
              = (5^2)(0.06)
              = 1.5 \mathrm{W}
      b.
            P = I^2 R
              = (5^2)(0.03)
              = 0.75 W
 11.
       W = VI\Delta t
      \Delta t = \frac{W}{VI}
          =\frac{\frac{3\ 000}{(200)(15)}}
           = 1 s
```

12. Power increases if the resistance decreases only when voltage is kept constant. Power increases if resistance increases, only when the current is kept constant.

Back to Unit 2: Assessment

SUBJECT OUTCOME VII ELECTRICITY AND ELECTRONICS: STATE AND APPLY PRINCIPLES AND COMPONENTS USED IN ELECTRONICS



Subject outcome

Subject outcome 4.2: State and apply principles and components used in electronics



Learning outcomes

- $\cdot\;$ Define capacitance and inductance and the application thereof.
- Describe semiconductors.
- Identify intrinsic properties and doping properties by design.
- Describe principles of the p-n junction.
- Identify basic principles of digital electronics logical gates.
- Identify symbols and use of active circuit elements and identify components from circuit diagram.
 - Range: Active circuit elements are LED, diode, transistor, and operational amplifier.



Unit 1 outcomes

By the end of this unit you will be able to:

- Define capacitance.
- Define inductance.
- Describe the applications of capacitance and inductance.



By the end of this unit you will be able to:

- Describe semiconductors.
- + Identify intrinsic properties and doping properties by design.

• Describe principles of the p-n junction.



Unit 3 outcomes

By the end of this unit you will be able to:

- Identify basic principles of digital electronics logical gates.
- Identify symbols and use of active circuit elements and identify components from circuit diagrams, namely:
 - LEDs
 - diodes
 - transistors
 - operational amplifiers.
Unit 1: Capacitance and inductance

LEIGH KLEYNHANS



- Define inductance.
- Describe the applications of capacitance and inductance.

What you should know

Before you start this unit, make sure you can:

- Understand the concept of electric field strength around charges and capacitance, and the calculation of capacitance. Refer to <u>level 3 subject outcome 4.1 unit 3</u> if you need help with this.
- Understand the concept of electromagnetic induction. Refer to <u>level 3 subject outcome 4.2 unit 2</u> if you need help with this.

Introduction

You have already learnt that a capacitor is a device used to store electrical charge and electrical energy. In this unit you will learn about capacitance, the amount of electrical charge and energy capacitors can store.

You have also learnt that induction is the process by which an emf is induced by a changing magnetic flux. In this unit, you will learn that the physical quantity that determines how effective a device is in doing this is called inductance.

Capacitance

Capacitance exists wherever conductive material is separated by insulating material. Capacitors can store energy as an electric field.



Figure 1: Circuit with a capacitor

The property of a capacitor to store charge on its plates in the form of an electrostatic field is called the capacitance of the capacitor. The capacitance of a capacitor is determined by the surface area of the conductive plates and the distance of separation between them. Altering any two of these values alters the value of its capacitance. Because capacitors store the energy of the electrons in the form of an electrical charge on the plates, the larger the plates and/or smaller their separation the greater the charge that the capacitor holds for any given voltage across its plates. In other words, larger plates, smaller distance, more capacitance.

The capacitance of a capacitor can be calculated using the following formula: $C = \frac{Q}{V}$.

Inductance

Inductance is a property of a current carrying conductor to build a magnetic field. Inductance indicates the amount of voltage that will be generated by an inductor because of changes in the rate of charge movement through the inductor.

All conductors, such as wires, have inductance. To create an inductor, we use techniques that strengthen a magnetic field and thereby increase inductance. A basic inductor is simply a coiled wire. Figure 2 shows how this structure concentrates magnetic field lines.



Figure 2: An induction coil

Inductance is the property of a device that tells us how effectively it induces an emf in another device. For a given coil, the ratio of the emf to the rate of change of the current in the coil is a measure of the inductance.

A large inductance may or may not be desirable. We want a transformer to have a large inductance, because its purpose is to induce an emf. But an appliance, such as an electric clothes dryer, can induce a dangerous emf on its metal case if the inductance between its coils and the case is large. One way to reduce inductance is to counter-wind coils to cancel the magnetic field produced.



Figure 3: The heating coils of an electric clothes dryer can be counter-wound so that their magnetic fields cancel one another

Applications of capacitance and inductance

Voltage sources provide the energy that allows an electrical circuit to carry out its intended function. However, a circuit is not simply divided into sources that supply energy and components that consume energy.

Capacitors and inductors are used to store energy and can function as temporary energy sources in circuits. Inductance and capacitance are independent properties of two different electrical components. While inductance is a property of a current carrying conductor to build a magnetic field, capacitance is a measure of the ability of a device to hold electrical charges.

A component's capacitance is a critical circuit-design parameter because it influences the rate at which voltage (or current) changes during charging and discharging. Higher capacitance means that the voltage across a capacitor will increase more slowly (when it is charging) and decrease more slowly (when it is discharging).

If a circuit with a capacitor is disconnected from a power supply, the capacitor will temporarily maintain voltage. If a circuit with an inductor is disconnected from a power supply, the inductor will temporarily maintain current. Another way of saying this is that capacitors 'resist' changes in voltage and inductors 'resist' changes in current.

Both these properties are used in various applications. For example, they are widely used in power networks, voltage-regulator circuits, and electronic circuit filters that regulate frequency signals.

Summary

In this unit you have learnt the following:

- Capacitance is the ability of a capacitor to store charge in the form of an electric field.
- Capacitance of a capacitor is dependent on the surface area of the conductive plates and the distance of separation between them.
- Inductance is the property of a device that determines how effectively it induces an emf in another device.
- Inductance is dependent on the ratio of the emf to the rate of change of the current in the coil.
- \cdot Capacitance is applied in electrical circuits to 'resist' changes in voltage.
- Inductance is applied in electrical circuits to 'resist' changes in current.
- Examples of the application of capacitance and inductance are in power networks, voltage-regulator circuits, and electronic circuit filters that regulate frequency signals.

Unit 1: Assessment

Suggested time to complete: 15 minutes

- 1. Define capacitance.
- 2. Define inductance.
- 3. Tabulate the differences between capacitance and inductance.
- 4. How are the properties of capacitance and inductance applied in electrical circuits?

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Unit 1: Assessment

- 1. Capacitance is the property of a current-carrying conductor to hold and store charge.
- 2. Inductance is the property of a current-carrying conductor to generates a magnetic field.
- 3.

Capacitance	Inductance
Property of a current-carrying conductor to hold and store charge	Property of a current-carrying conductor to generate a magnetic field
Based on creating an electric field	Based on a magnetic field
Associated with electrical components called capacitors which are parallel plates separated by a dielectric	Associated with electrical components called inductors which are coiled current-carrying wires
'Resist' changes in voltage	'Resist' changes in current

4. Capacitance and inductance are used in power networks, voltage-regulator circuits, and electronic circuit filters that regulate frequency signals.

Back to Unit 1: Assessment

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Unit 2: Semi-conductors

LEIGH KLEYNHANS



By the end of this unit you will be able to:

- Describe semiconductors.
- Identify intrinsic properties and doping properties by design.
- Describe principles of the *p*-*n* junction.

What you should know

Before you start this unit, make sure you can:

- Explain the structure of the atom. Refer to level 2 subject outcome 5.2 unit 1 if you need help with this.
- Work with the properties of elements and their arrangement in the periodic table. Refer to <u>level 2</u> <u>subject outcome 5.3 units 1, 2</u> and <u>3</u> if you need help with this.
- Define the principles of bonding. Refer to level 2 subject outcome 5.4 unit 3 if you need help with this.

Introduction

The classification of substances would be easy if they could be divided simply into conductors and insulators. However, if this was the case, we wouldn't be able to make computers, CD players, cell phones, iPods and many other things which are part of modern life. The operation of modern technological devices is based on substances that are classified as semiconductors. Semiconductors are defined by their unique electric conductive behaviour, somewhere between that of a conductor and an insulator.

Atomic structure of semiconductors

To understand how semiconductors work, you must understand how electrons are organised in an atom. You should remember from previous physical science levels that the electrons in an atom are organised in shells or energy levels. The outermost shell is called the valence shell.

Metals tend to be good conductors of electricity because they usually have 'free electrons' in their valence shells that can move easily between atoms, and electricity involves the flow of electrons.

In insulators, the electrons in the valence shell are the ones that form bonds with neighbouring atoms. These bonds are called covalent bonds, and the electrons cannot move freely.

Carbon, silicon, and germanium have a unique property in their electron structure – each has four electrons in its outer shell. It is possible for all the valence electrons to be shared between adjacent atoms and the atoms arrange themselves into structures called crystals. Semiconductors are made of these crystals, usually silicon crystals.

In figure 1, each circle represents a silicon atom, and the lines between the atoms represent the shared

electrons. Each of the four valence electrons in each silicon atom is shared with one adjacent silicon atom. Thus, each silicon atom is bonded with four other silicon atoms.



Pure silicon crystals are not useful electronically. There are no 'free electrons', so it is an insulator. However, under certain conditions the crystal can conduct electricity.

Extrinsic semiconductors

The process of deliberately introducing other elements into a crystal is called doping. The element introduced by doping is called a dopant. By carefully controlling the doping process and the dopants that are used, silicon crystals can transform into one of two distinct types of conductors:

- n-type semiconductor
- p-type semiconductors.

An **n-type semiconductor** is created when the dopant is an element that has five electrons in its valence layer. Phosphorus is commonly used for this purpose. The phosphorus atoms join the crystal structure of the silicon, each one bonding with four adjacent silicon atoms just like a silicon atom would. Because the phosphorus atom has five electrons in its valence shell, but only four of them are bonded to adjacent silicon atoms, the fifth valence electron is left unbonded. These extra valence electrons in the phosphorous atoms start to behave like the 'free' valence electrons in a regular conductor such as copper and the doped silicon can conduct electricity. Because this type of semiconductor has extra electrons (which are negatively charged), it is called an n-type semiconductor.



Figure 2: Silicon doped with phosphorus to make an n-type semiconductor

A **p-type semiconductor** is created when the dopant (such as boron) has only three electrons in the valence shell. When a small amount is incorporated into the crystal, the atom can bond with four silicon atoms, but since it has only three electrons to offer, a hole is created. The hole is a vacant 'space' which behaves like a positive charge (attracting electrons to it), so semiconductors doped in this way are called p-type semiconductors.

Like a positive charge, holes attract electrons. But when an electron moves into a hole, the electron leaves a new hole at its previous location. Thus, in a p-type semiconductor, holes are constantly moving around within the crystal as electrons constantly try to fill them up.



Figure 3: Silicon doped with boron to make a p-type semiconductor

A minute amount of either n-type or p-type doping turns a silicon crystal from a good insulator into a viable (but not great) conductor – hence the name 'semiconductor'. When voltage is applied to either an n-type or a p-type semiconductor, current flows, for the same reason that it flows in a regular conductor: the negative side of the voltage pushes electrons, and the positive side pulls them. The result is that the random electron

and hole movement that is always present in a doped semiconductor becomes organised in one direction, creating measurable electric current.

Principles of p-n junctions

N-type and p-type silicon are not that amazing by themselves; but when you put them together, you get some interesting behaviour at the junction. This arrangement is called a p-n junction and is the basis of electronic components called diodes.

A diode is the simplest possible semiconductor device. A diode allows current to flow in one direction but not the other. You may have seen turnstiles at a stadium or a station that let people go through in only one direction. A diode is a one-way turnstile for electrons. When you put n-type and p-type silicon together as shown in figure 4, you get an interesting phenomenon that gives a diode its unique properties.



Even though n-type silicon by itself is a conductor, and p-type silicon by itself is also a conductor, the combination shown in figure 4 does not conduct any electricity. The negative electrons in the n-type silicon get attracted to the positive terminal of the battery. The positive holes in the p-type silicon get attracted to the negative terminal of the battery. No current flows across the junction because the holes and the electrons are each moving in the wrong direction. This arrangement is referred to as reverse biased. If you flip the battery around, the diode will conduct electricity. The free electrons in the n-type silicon are repelled by the negative terminal of the battery.



Figure 5: A p-n junction with the n-type connected to the negative terminal of the battery and the p-type connected to the positive terminal.

The holes in the p-type silicon are repelled by the positive terminal. At the junction between the p-type and n-type silicon, holes and free electrons meet. The electrons fill the holes. Those holes and free electrons cease to exist, and new holes and electrons spring up to take their place. The effect is that current flows through the junction. This arrangement is referred to as forward biased. This property of diodes, which allows current to flow only if they are placed in a circuit in a forward biased position, make them the electronic version of a valve. The application of this property in electronic devices will be covered in the next unit.

Intrinsic semiconductors

Intrinsic semiconductors do not contain any impurities. They will be insulators as there are no free electrons to conduct electricity. However, if the substance is heated, electrons that are involved in covalent bonds can become 'excited' and free themselves from the bond. This creates free electrons (negative charge carriers) and creates holes in the position they left (positive charged areas). In this way a pure semiconductor can start to conduct electricity when connected to a voltage source.



Figure 6: Silicon behaving as an intrinsic semiconductor when heated

The semiconductor will only conduct electricity when heated and will become an insulator when it cools down.

Summary

In this unit you have learnt the following:

- Semiconductors are elements that have four valence electrons forming covalent bonds between the atoms forming a crystal.
- Extrinsic semiconductors are formed by introducing small amounts of another element in a process called doping. This allows it to conduct electricity.
- If a dopant with five valence electrons is used, the semiconductor is called an n-type semiconductor.
- If a dopant with three valence electrons is used, the semiconductor is called a p-type semiconductor.
- Diodes are constructed by joining a p-type semiconductor to an n-type semiconductor called a p-n junction.
- Diodes allow current to flow in a circuit only when they are connected in the forward biased direction.
- Intrinsic semiconductors can only conduct electricity when they have been heated and electrons in covalent bonds become free, leaving behind positive holes.

Unit 2: Assessment

Suggested time to complete: 20 minutes

- 1. Explain why elements that form semiconductors have a crystalline structure.
- 2. Name two substances that are classified as semiconductors.
- 3. Give one word for the following:
 - a. the process of adding impurities to create a semiconductor.
 - b. substances that can be used for the process in question 3 a).
- 4. Explain the difference between an n-type and a p-type semiconductor.
- 5. State whether the following would be an n-type or p-type semiconductor:
 - a. germanium doped with boron
 - b. silicon doped with phosphorus
 - c. germanium doped with aluminium
 - d. silicon doped with arsenic
- 6.
- a. Name the electronic component that consists of a p-n junction.
- b. State how this component needs to be connected in a circuit to allow current to flow.
- 7. Explain how intrinsic semiconductors allow current to flow.

The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Unit 2: Assessment

- 1. They have four valence electrons so will form four covalent bonds with surrounding atoms in a symmetrical lattice pattern.
- 2. silicon, germanium.
- 3.
- a. doping
- b. dopant
- 4. An n-type semiconductor has been doped with an element with five valence electrons creating 'free electrons' not involved in covalent bonds with surrounding atoms. These electrons will be the charge carriers when connected to a voltage source.

A p-type semiconductor has been doped with an element with three valence electrons, creating positive holes where there is an unpaired electron. These positive holes act as charge carriers as they attract electrons towards them when connected to a voltage source.

- 5.
- a. p-type (boron has three valence electrons)
- b. n-type (phosphorus has five valence electrons)
- c. p-type (aluminium has three valence electrons)
- d. n-type (arsenic has five valence electrons)

- 6.
- a. diode
- b. The side of the diode with the p-type semiconductor must be connected to the positive terminal of the power source and the side with the n-type semiconductor must be connected to the negative terminal of the power source. This is forward biased.
- 7. Intrinsic semiconductors do not have free electrons but heating them will cause electrons in covalent bonds to become free. They can then become charge carriers at this higher temperature and conduct electricity.

Back to Unit 2: Assessment

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Unit 3: Digital electronics

LEIGH KLEYNHANS



What you should know

Before you start this unit, make sure you can:

• Describe the principles of a p-n junction in diodes. Refer to <u>level 4 subject outcome 4.2 unit 2</u> if you need help with this.

Introduction

Circuits often do not only consist of a power source, conducting wires, resistors, capacitors, and inductors. In this unit you will be introduced to further circuit components, specifically those used in electronic or digital circuits.

What are logic gates?

Logic gates are the building blocks of digital systems. They are basically used to perform logical operations by acting as a switch between options. A logic gate is an electronic device that creates logical decisions depending on the various combinations of digital signals accessible on its inputs (sensors). A digital logic gate can have more than one input signal but has only one digital output signal.

A logic gate accepts the input and operates on a required condition. If a certain condition is true, it turns 'ON' (also referred to as HIGH) and goes 'OFF' (also referred to as LOW) when the condition is false.

The functioning of a logic gate is expressed using what is referred to as a truth table. It displays all possible combinations of the inputs and a column against each row shows the corresponding output value. Figure 1 shows an example of logic gate circuit which has three inputs A, B, C, and single output Y.



Figure 1: Example of logic gates using symbols

Types of logic gates

There are seven types of logic gates. They are:

- NOT gate (inverter)
- AND gate (multiplication)
- OR gate (addition)
- NAND gate
- NOR gate
- XOR gate
- XNOR gate.

NOT gates (inverter)

The output of a NOT gate is 'high' if the input is 'low' and vice versa.



Α	Y
0	1
1	0

Figure 2: Symbol and truth table of a NOT gate

AND gates (multiplication)

The output of an AND gate is 'high' if both the inputs are high, and the output is 'low' if both or either of the inputs is 'low'.



Figure 3: Symbol and truth table of an AND gate

OR gates (addition)

The output of an OR gate is 'high' if both the inputs are 'high' or one of the inputs is 'high'. The output is 'low' if both the inputs are 'low'.



Figure 4: Symbol and truth table of an OR gate

NAND gates

The output of a NAND gate is 'high' if both the inputs are 'low' and if either of the inputs is 'low'. The output is 'low' if both the inputs are 'high'.



Figure 5: Symbol and truth table of a NAND gate

NOR gates

The output of a NOR gate is 'high' if both the inputs are 'low'. The output is 'low' if both the inputs are 'high' or either of the inputs is 'high'.



Figure 6: Symbol and truth table of a NOR gate

XOR gates

The output of a XOR gate is 'high' if either of the inputs is 'high'. The output is 'low' if both the inputs are 'high' or if both the inputs are 'low'.



Figure 7: Symbol and truth table of a XOR gate

XNOR gates

The output of a XNOR gate is 'high' if both the inputs are 'high' or if both the inputs are 'low'. The output is 'low' if either of the input is 'low'.



Figure 8: Symbol and truth table of a XNOR gate

The working principle of logic gates

To understand the working principle of logic gates, let us consider the example of a street lighting system. Figure 9 shows the circuit diagram of a street lighting system which consists of logic gates (a NOT gate and an OR gate), a switch and a light sensor. The operational condition for the streetlights to function is as follows:

Switch: On = 1, Off = 0 Light sensor: Dark = 0, Bright = 1 Streetlight: On = 1, Off = 0

Based on this, we can prepare a truth table as shown in figure 9:



Figure 9: A circuit diagram of street lighting system and truth table

Applications of logic gates

There are a number of applications of logic gates.

- They are used in circuits involving computation and processing.
- They are also used in push button switches, e.g. a door bell.
- They are used in the functioning of streetlights.
- AND gates are used to enable/inhibit data transfer.
- NAND gates are used in burglar alarms and buzzers.



Active circuit elements

The components you have been learning about so far — resistors, capacitors, and inductors — are called *passive* components. They do not change their behaviour and therefore always have the same response to changes in voltage or current. *Active* components are quite different. They respond to changes in input, and this allows them to be used as circuit components called diodes, transistors, and amplifiers, which are used in calculators and computers.

Diodes

A diode is an electronic device that allows current to flow in one direction only.



Figure 10: A diode circuit symbol

A diode consists of two doped semiconductors joined together so that the resistance is low when connected one way and very high the other way.

Look at the circuit diagrams in figure 11. In the circuit diagram on the left, the diode is forward biased, and current is permitted. The negative terminal of the battery is connected to the negative terminal of the diode. In the circuit diagram on the right, the diode is reverse biased and current flow is not allowed. The negative terminal of the battery is connected to the positive terminal of the diode.



Figure 11: Circuit diagrams for forward and reverse biased diode connections

Diodes are used for:

- power conversion to convert AC power to DC power.
- demodulation of signals to remove the negative component of an AC signal.
- · over-voltage protections protection devices for sensitive electronic components.
- current steering to make sure current only flows in the proper direction.

The light-emitting diode (LED)

A light-emitting diode (LED) is a diode device that emits light when charge flows in the correct direction through it.



Figure 12: The symbol for a light-emitting diode

The colour depends on the semiconducting material used to construct the LED, and can be in the nearultraviolet, visible, or infrared part of the electromagnetic spectrum.

The wavelength of the light emitted, and therefore its colour, depends on the materials forming the pn junction. A normal diode, typically made of silicon or germanium, emits invisible far-infrared light (so it cannot be seen by the human eye), but the materials used for an LED can emit light corresponding to nearinfrared, visible or near-ultraviolet frequencies.

LEDs have many uses, such as for:

- thin, lightweight message displays, e.g. in public information signs (at airports and railway stations, among other places)
- status indicators, e.g. on/off lights on professional instruments and consumers audio/video equipment
- infrared LEDs in remote controls (for TVs, VCRs, etc.)
- traffic signals; clusters of LEDs are used in replacing ordinary bulbs behind coloured glass
- car indicator lights and bicycle lighting
- calculator and measurement instrument displays (although now mostly replaced by LCDs)
- indicator and alphanumeric displays in environments where night vision must be retained (these use red and yellow LEDs): aircraft cockpits, submarine and ship bridges, astronomy observatories, and in the field, e.g. night time animal watching and military field use
- in photographic darkrooms, for providing lighting which does not lead to unwanted exposure of the film (these use red and yellow LEDs)
- illumination, e.g. torches and backlighting for LCD screens
- signalling/emergency beacons and strobes
- movement sensors, e.g. in a computer mouse.

The typical working lifetime of LEDS is ten years, which is much longer than the lifetime of most other light sources. LEDs fail by dimming over time, rather than the abrupt burn-out of incandescent bulbs. LEDs give off less heat than incandescent light bulbs and are less fragile than fluorescent lamps.

Because they are monochromatic, LED lights have great power advantages over white lights in cases where a specific colour is required. Unlike the white lights, the LED does not need a filter that absorbs most of the emitted white light. LED lights only emit one colour of light and are available in a wide range of colours.

Transistors

The diode is the simplest semiconductor device, made up of a p-type semiconductor and an n-type semiconductor in contact. It can conduct in only one direction, but it cannot control the size of an electric current. Transistors are more complicated electronic components which can control the size of the electric current flowing through them.

A small current can be used to turn the transistor on and off. The transistor then controls a more complicated or powerful current through other components. When a transistor is used in this way it is said to be in *switching* mode as it is acting as a remotely controlled switch. Switching circuits can be used in a computer to process and store digital information. A computer would not work without the millions (or billions) of transistors in it.



The transistor is considered by many to be one of the greatest discoveries or inventions in modern history. Key to the importance of the transistor in modern society is its ability to be produced in huge numbers using simple techniques, resulting in their low cost. Computer 'chips' consist of millions of transistors. The low cost has meant that the transistor has become an almost universal tool for non-mechanical tasks. Whereas a common appliance, say a refrigerator, would have used a mechanical device for control, today it is often less expensive to simply use a few million transistors and the appropriate computer program to carry out the same task. Today transistors have replaced almost all electromechanical devices, most simple feedback systems, and appear in huge numbers in everything from computers to cars.

The operational amplifier

The operational amplifier is a special kind of electronic component which is made from transistors and is used to change voltage. Operational amplifiers are usually called op-amps for short. They are used extensively in all kinds of audio equipment (amplifiers, mixers, etc.) and in instrumentation. They also have many other uses in other circuits, for example comparing voltages from sensors.



Figure 14: The circuit symbol for an operational amplifier

The operational amplifier has two input terminals and one output terminal. The voltage of the output terminal is proportional to the difference in voltage between the two input terminals. The output terminal is on the right (at the sharp point of the triangle). The two input terminals are drawn on the left. One input terminal (labelled with a + on diagrams) is called the **non-inverting input**. The other input terminal (labelled -) is called the **inverting input**. The labels + and – have nothing to do with the way in which the operational amplifier is connected to the power supply. Operational amplifiers must be connected to the power supply, but this is taken for granted when circuit diagrams are drawn, and these connections are not shown on circuit diagrams. Usually, when drawing electronic circuits, '0V' is taken to mean the negative terminal of the power supply. This is not the case with op-amps. For an op-amp, '0V' refers to the voltage midway between the + and – of the supply.



- 1. What is a diode?
- 2. What is a diode made of?
- 3. What is the term which means that a diode is connected the 'wrong way' and little current is flowing?
- 4. What is an LED?
- 5. List five applications of LEDs.
- 6. What are transistors used for?
- 7. What are operational amplifiers used for?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- A logic gate is an electronic device that creates logical decisions depending on the various combinations of digital signals accessible on its inputs (sensors).
- The functioning of a logic gate is in a truth table which displays all possible combination of the inputs and corresponding output value.
- Logic gates are used in computers, push button switches, controlling streetlights, data transfer, burglar alarms and buzzers.
- Active circuit components respond to changes in voltage and current. They include diodes, light emitting diodes, transistors, and operational amplifiers.
- A diode is an electronic device that allows current to flow in one direction only.
- Diodes are used to convert AC power to DC power, to remove the negative component of an AC signal, to protect sensitive electronic components and to make sure current only flows in the proper direction.
- A light-emitting diode (LED) is a diode device that emits light when charge flows in the correct direction through it.
- LEDs are used for signs, status indicators, remote controls, traffic lights, lighting, and movement sensors.
- Transistors are electronic components which can control the size of the electric current flowing through them.
- Transistors act as remote-controlled switches.
- The operational amplifier is an electronic component which is made from transistors and is used to change voltage.
- Operational amplifiers are used extensively in audio equipment (amplifiers, mixers etc.) and in instrumentation.

Unit 3: Assessment

Suggested time to complete: 30 minutes

1. Write the number of the correct description for the logic gate next to the letter of its name:

Logic gate	Description	
a. NAND	 The output is 'high' if either of the inputs is 'high'. The output is 'low' if both the inputs are 'high' or if both the inputs are 'low'. 	
b. NOT	The output is 'high' if both the inputs are 'high', and the output is 'low' if both or either of the inputs is 'low'.	
c. OR	 The output is 'high' if both the inputs are 'high' or if both the inputs are 'low'. The output is 'low' if either of the input is 'low'. 	
d. AND	4. The output is 'high' if both the inputs are 'low' and if either of the inputs is 'low'. The output is 'low' if both the inputs are 'high'.	
e. XOR	The output is 'high' if both the inputs are 'low'. The output is 'low' if both the inputs are 'high' or either of the inputs is 'high'.	
f. XNOR	 The output is 'high' if both the inputs are 'high' or one of the inputs is 'high'. The output is 'low' if both the inputs are 'low'. 	
g. NOR	7. The output is 'high' if the input is 'low' and vice versa.	

- 2. State the difference between passive and active circuit components.
- 3. Tabulate the active circuit elements you have learnt under the headings: name, symbol, use.
- 4. Which of the light bulbs in the circuits below will light up? Explain.



5. A LED is a diode that emits light, and for these special types of diodes there are many applications. The digital number display on Izaan's calculator is produced by a system of LEDs. The circuit diagram for the LEDs for a single number 8 is given beneath.



- a. What switches must be opened to change this to a number 2?
- b. What number would be seen if all switches were closed but LED "c" is connected in reverse? Explain briefly.

The <u>full solutions</u> are at the end of the unit.

Unit 3: Solutions

Exercise 3.1

- 1. A logic gate is an electronic device that creates logical decisions depending on the various combinations of digital signals accessible on its inputs.
- 2. There are seven types of logic gates. They are:

- NOT gate (inverter)
- AND gate (multiplication)
- OR gate (addition)
- NAND gate
- NOR gate
- XOR gate
- XNOR gate.
- 3. Logic gates are used in computers, push button switches, controlling streetlights, data transfer, burglar alarms and buzzers.

Back to Exercise 3.1

Exercise 3.2

- 1. A diode is an electronic device that allows current to flow in one direction only.
- 2. Diodes consist of a p-type and an n-type semiconductor in a p-n junction.
- 3. Reverse biased.
- 4. Light emitting diode.
- 5. signs status indicators remote controls traffic lights lighting movement sensors.
- 6. Transistors are used as remote-controlled switches.
- 7. Operational amplifiers are used in audio equipment such as amplifiers and mixers, and in instrumentation.

Back to Exercise 3.2

Unit 3: Assessment

1.

- a. 4
- b. 7
- c. 6
- d. 2
- e. 1
- f. 3
- g. 5
- 2. **Passive** components do not change their behaviour and therefore always have the same response to changes in voltage or current. **Active** components respond to changes in voltage and current.
- 3. Table:





- 4. B1 the diode is connected in the forward biased direction, the cathode (positive) is connected to the positive terminal of the battery and the anode (negative) is connected to the negative terminal of the battery.
- 5.
- a. Switches b) and e)
- b. 9

Back to Unit 3: Assessment

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SUBJECT OUTCOME VIII MATTER AND MATERIALS: IDENTIFY AND CRITICALLY EVALUATE THE IMPACT OF ATOMIC NUCLEI ON THE QUALITY OF HUMAN, ENVIRONMENTAL AND SOCIO-ECONOMIC DEVELOPMENT

Subject outcome

Subject outcome 5.1: Identify and critically evaluate the impact of atomic nuclei on the quality of human, environmental and socio-economic development



Learning outcomes

- Describe the nuclear structure of an atom and its stability.
- Define radioactivity and identify its particles and how it is measured.
- Identify and evaluate examples of use in industry and the biological effects of radiation and safety.
- Differentiate between nuclear fusion and nuclear fission and evaluate nuclear power.



Unit 1 outcomes

By the end of this unit you will be able to:

- Describe and define the nuclear structure of an atom and its stability.
- Explain radioactivity and how it is measured.
- Understand nuclear fusion and fission and nuclear power.

Unit 1: Radioactivity

EMMA HARRAGE



- Describe and define the nuclear structure of an atom and its stability.
- Explain radioactivity and how it is measured.
- Understand nuclear fusion and fission and nuclear power.

What you should know

Before you start this unit, make sure you:

• Understand isotopes. Refer to level 2 subject outcome 5.2 unit 2 if you need help with this.

Introduction¹

In this unit you will learn about radioactivity. Radioactivity occurs when a nucleus in an atom is unstable because of the number of protons and neutrons. We are exposed to radiation every day from the sun – the sun is a giant ball of gases which produces light and heat because of nuclear fusion. Radioactive elements such as uranium are used to generate electricity. Over exposure to radiation can cause you to develop radiation sickness and cancers.

Atoms and stability

As we learnt in <u>level 2 subject outcome 5.2 unit 2</u>, atoms are composed of subatomic particles – protons, neutrons, and electrons. Protons and neutrons are in the nucleus and provide most of the mass of the atom, while electrons circle the nucleus in shells and subshells and account for an atom's size. Remember, the notation for representing an isotope of a particular atom is as follows:

1. Parts of the text in this unit were sourced from <u>chem.libretexts.org</u>: <u>Introductory</u> <u>Chemistry</u>, released under a CC BY NC SA licence.

Figure 1: An isotope of the element carbon

Mass number/	
number of protons	
and neutrons	12
Atomic number/number of	6 U
protons	

Number of neutrons = mass number – atomic number So 12 - 6 = 6. Carbon has 6 neutrons

An atom is stable because of a balanced nucleus that does not contain excess energy. If the forces between the protons and the neutrons in the nucleus are unbalanced, then the atom is unstable. Stable atoms retain their form indefinitely, while unstable atoms undergo radioactive decay. Most naturally occurring atoms are stable and do not decay.

Unstable atoms are radioactive and decay after a certain amount of time. Different unstable atoms decay through different processes, such as the discharge of a proton or a neutron or the conversion of a proton to a neutron or a neutron to a proton. This result causes the atom to remain unstable until it decays to form a new stable atom.

Most elements have isotopes; in many cases one or more of these isotopes is radioactive, whilst others are not radioactive. For this reason, it is usual to refer to individual isotopes when describing radioactivity. e.g. cobalt-60 is radioactive, but cobalt-59 is not.

Major types of radioactivity

Atomic theory in the 19th century presumed that nuclei had fixed compositions. But in 1896, the French scientist Henri Becquerel found that a uranium compound placed near a photographic plate made an image on the plate, even if the compound was wrapped in black cloth. He reasoned that the uranium compound was emitting some kind of radiation that passed through the cloth to expose the photographic plate. Further investigations showed that the radiation was a combination of particles and electromagnetic rays, from the atomic nucleus. These rays were collectively called radioactivity. There are three main forms of radioactive emissions.

Alpha (α) particles

The first is called an alpha particle, which is symbolised by the Greek letter α . An alpha particle is composed of two protons and two neutrons, and so it is the same as a helium nucleus: $\frac{4}{2}$ He.

When a radioactive atom emits an alpha particle, the original atom's atomic number decreases by two (because of the loss of two protons), and its mass number decreases by four (because of the loss of 2 protons and 2 neutrons). We can represent the emission of an alpha particle with a chemical equation. For example, the alpha particle emission of uranium, 235 U, is as follows:

 $^{235}_{92}{
m U}
ightarrow ^4_2{
m He} + ~^{231}_{90}{
m Th}$

How do we know that a product of the reaction is $^{231}_{90}$ Th?

We use the law of conservation of matter, which says that matter cannot be created or destroyed. This means there must be the same number of protons and neutrons on both sides of the chemical equation.

If the uranium nucleus loses 2 protons, there are 90 protons remaining, identifying the element as thorium. If we lose 4 nuclear particles of the original 235, there are 231 remaining. Thus, we use subtraction to identify the isotope of the thorium atom, $^{231}_{90}$ Th.

When one element changes into another in this manner, it undergoes radioactive decay.

Radon has an atomic number of 86, so the isotope is represented as $^{222}_{86}$ Rn. We represent the alpha particle as $^{4}_{2}$ He and use subtraction:

222 - 4 = 21886 - 2 = 84

We can now identify the decayed isotope as an isotope of polonium, $^{218}_{84}$ Po:

 $^{222}_{86}\mathrm{Rn} \rightarrow^{4}_{2}\mathrm{He} +^{218}_{84}\mathrm{Po}$

Beta (β) particles

The second major type of radioactive emission is called a beta particle, symbolised by the Greek letter β . A beta particle is an electron ejected from the nucleus (not from the shells of electrons about the nucleus). We can also represent a beta particle as ${}^{0}_{-1}e$ or β -. The net effect of beta particle emission on a nucleus is that a neutron is converted to a proton. The overall mass number stays the same, but because the number of protons increases by one, the atomic number goes up by one. Carbon-14 decays by emitting a beta particle:

 ${}^{14}_6{
m C}
ightarrow {}^{14}_7{
m N} + {}^0_{-1}e$

Again, the sum of the atomic numbers is the same on both sides of the equation, as is the sum of the mass numbers. Note that the electron is assigned an 'atomic number' of -1, equal to its charge.

The emission of an α -particle or a β -particle from a nucleus is an example of a nuclear reaction. A nuclear reaction results in the change in the composition of a nucleus and therefore results in the formation of new atoms with different atomic numbers. Chemical reactions, by contrast, do not change the composition of a nucleus – they only involve the rearrangement of the electrons in orbitals and therefore do not result in the formation of new atoms.

For this reason, nuclear equations always include atomic numbers and mass numbers of every particle, as they change during the reaction.

Gamma (γ) rays

The third major type of radioactive emission is not a particle but rather a very energetic form of electromagnetic radiation called gamma rays, symbolised by the Greek letter γ . Gamma rays themselves do not carry an overall electrical charge, but they may knock electrons out of atoms in a sample of matter and make it electrically charged.

For example, in the radioactive decay of radon, $^{222}R_a$, both alpha and gamma radiation are emitted, with the latter having an energy of 8.2×10^{-14} J per nucleus decayed. This may not seem like much energy, but if 1 mol of radon atoms were to decay, the gamma ray energy would be 49 million kJ (which is the same amount of energy in 82 000 bananas!).

$^{222}_{86}\mathrm{Rn} ightarrow^{218}_{84} \mathrm{Po} + ^4_2 \mathrm{He} + \gamma$

The emission of an α - or β -particle often results in a much more stable nucleus with a much lower potential

energy; the potential energy is converted into a high-energy photon of electromagnetic gamma radiation. Gamma rays are only emitted at the same time as α - or β -particles.



Figure 2: α -particles, β -particles and γ -rays are collectively known as 'nuclear radiation'

Туре	Symbol	Radiation emitted	Description
Alpha	$^4_2\mathrm{He}~\mathrm{or}~^4_2lpha$	$\frac{4}{2}\alpha$	High energy helium nuclei consisting of two protons and two neutrons. It has a large mass, compared to other ionising radiations, and a strong positive charge.
Beta	${0 \over -1}{ m e} { m or} {0 \over -1} eta$	$^{0}_{-1}eta$	High energy electrons. It has a very small mass and a negative charge.
Gamma	γ	${}^0_0\gamma$	Very high energy electromagnetic radiation. Gamma rays are caused by changes within the nucleus. They are part of the electromagnetic spectrum and so travel at the speed of light. They have no mass and no charge.

Below is a summary of the three main forms of radioactive emissions



Exercise 1.1

1.

- a. Describe an alpha particle; explain what happens to a nucleus after it emits an alpha particle.
- b. Describe a beta particle; explain what happens to a nucleus after it emits a beta particle.
- c. Explain why gamma rays are sometimes emitted at the same time as an alpha or beta particle
- 2. Write the nuclear equation that represents the radioactive decay of polonium-209 by alpha particle emission and identify the decayed isotope.

The <u>full solutions</u> are at the end of the unit.

Properties of radiation

Alpha particles collide easily with other particles, pulling electrons away from them until the α -particle has gained two electrons and become a stable helium atom. As a result α -particles are very strongly ionising – any particles they hit are likely to lose electrons and become ionised. Because of this, α -particles are very quickly destroyed, and they generally travel no further than 4 cm in air and are easily stopped by a thin piece of paper.

Beta particles are much smaller and so collide less easily with other particles. When they do collide with other particles, they transfer energy to them, and this can result in other electrons being lost from that atom. As a result β -particles are also ionising, although much less so than α -particles; eventually the β -particle will slow down and be absorbed into the electron shells of another atom. β -particles have more penetrating power than α -particles; they can travel a long distance through air and can pass through paper but can be stopped by a thin sheet of metal.

Gamma rays have no charge and no mass, they therefore do not cause ionisation in other particles and often pass through them completely without being absorbed. As a result, γ -rays have a very low ionising power but a very high penetrating power. They cannot be completely stopped and a few centimetres of lead or several metres of concrete is needed to significantly reduce their penetration.



Detecting radiation

The ionising properties of radiation mean that it can be detected by a piece of equipment known as a Geiger Muller tube. This tube contains gaseous particles which become ionised by the radiation, creating a current which can be detected; each radioactive particle or ray produces one pulse. The total concentration

of radiation in a particular location can be measured from the number of pulses per second; a Geiger-Muller tube does not distinguish between alpha, beta, and gamma radiation.



Figure 4: A Geiger Müller tube is used to detect levels of radiation

All three types of radiation are high in energy and if living cells are exposed to significant quantities of radiation, they can be seriously damaged. Sometimes the cells are killed and sometimes the cells will mutate and become cancerous; either way, high levels of exposure to radiation can be fatal.

The relative dangers of alpha, beta and gamma radiation are directly linked to their ionising power, their penetrating power and how the exposure takes place:

- Alpha and beta radiation is highly ionising but cannot penetrate skin; external alpha and beta radiation
 is therefore not considered dangerous unless it is present in large quantities and very close, in which
 case it will kill skin cells (burn the skin). If radioactive atoms are ingested, injected, or inhaled, however,
 they can be very dangerous as they release ionising radiation inside the body; this can kill healthy cells
 or turn them into cancerous cells.
- Gamma radiation can pass through the body due to its low ionising power; small quantities of gamma radiation are not considered dangerous. In fact, small quantities of gamma radiation can come from space and are emitted by rocks and soil. However, large quantities of gamma radiation can be dangerous wherever it is, and you can be harmed or killed by gamma radiation without inhaling, injecting, or ingesting radioactive material.

Radioactive decay

The rate of chemical reactions depends on various factors including temperature, pressure or concentration, surface area and the presence of a catalyst. The rate at which a nucleus emits α -particles and β -particles does not depend on temperature, surface area or any catalyst, it depends on how many radioactive atoms are present in the sample. Nuclear decay reactions occur spontaneously under all conditions and produce more stable daughter nuclei, whereas nuclear transmutation reactions are induced and form a product nucleus that is more massive than the starting material.

The rate of radioactive decay of a particular isotope is directly proportional to the number of atoms of that isotope in the sample.

As a result of this, the time taken for half of a sample radioactive isotope to decay is fixed for every
radioactive isotope. The time taken for half of the atoms to decay is called the half-life of that isotope and the half-life of radioactive isotopes can vary from fractions of a second to millions of years.



Figure 5: the radioactive decay of an isotope from the 'parent' isotope to the 'daughter' isotope

Isotope	Half-life			
⁹⁰ Kr	33 s			
$^{108}\mathrm{Ag}$	$2.4\mathrm{min}$			
131 I	8 days			
$^{14}\mathrm{C}$	5730 years			
$^{238}\mathrm{U}$	4.46 billion years			

Table 1: Examples of half-lives of certain radioactive isotopes

After one half-life, 50% of a sample will have decayed and 50% will be remaining; after two half-lives, 75% of a sample will have decayed and 25% will be remaining; after three half-lives, 87.5% of a sample will have decayed and 12.5% will be remaining.



Figure 6: The half-life of 10 grams of cobalt-60





Example 1.2

 60 Co has a half-life of 5.3 years; starting with 1000 atoms, how many atoms will have decayed after 10.6 years and how many will remain?

Solution

10.6 years = $\frac{10.6}{5.3}$ = 2 half-lives, so 75% or 750 atoms will have decayed and 25% or 250 atoms will be remaining.



Example 1.3

A sample of Ac -225 originally contained 8.0 ug. After 720 hours, how much of the original Ac -225 remains? The half-life of this isotope is 10 days.

Solution

To determine the number of half-lives (n), both time units must be the same:

*20*l*720 hours $\times \frac{1 \text{ day}}{24 \text{ hours}} = 30 \text{ days}$ n = 3 = $\frac{30 \text{ days}}{10 \text{ days}}$

How much mass remains = $\frac{1}{2^3}(8.0 \text{ ug})$

After 720 hours, 1.0 ug of the material remains as Ac –225.

The more stable a nucleus, the longer its half-life; the less stable a nucleus, the shorter its half-life. Nuclei which are completely stable (i.e. non-radioactive) do not have a half-life and have an infinite lifetime.

Radioactive isotopes with a long half-life are considered more of an environmental hazard than those with a short half-life, as they remain active in the environment for a much longer time.



2. 108 Ag has a half-life of 2.4 minutes; calculate the percentage of a sample which will have decayed after 7.2 minutes.

The <u>full solutions</u> can be found at the end of the unit.

Nuclear reactions and nuclear energy

The emission of an α -particle or a β -particle from a nucleus (radioactive decay) is just one example of a nuclear reaction; there are two other important types of nuclear reactions, nuclear fission and nuclear fusion.

Nuclear fission

Nuclear fission is the break-up of a large nucleus to form two or more smaller nuclei; it does not occur spontaneously and only happens when a nucleus is bombarded with a neutron.



The two smaller nuclei are usually more stable than the original large nucleus, so a lot of heat energy and gamma rays are released during nuclear fission; in addition, the neutrons produced can collide with more 235 U atoms and cause further fission, starting a nuclear chain reaction.

During nuclear fission, the nucleus usually divides asymmetrically rather than into two equal parts, as shown in figure 7. In addition, every fission event of a given nuclide does not give the same products; more than 50 different fission modes have been identified for uranium-235, for example. Consequently, nuclear fission of a fissile nuclide can never be described by a single equation.



As shown in figure 8, the collision of a relatively slow-moving neutron with a fissile nucleus can split it into two smaller nuclei with the same or different masses. Neutrons are also released in the process, along with a great deal of energy.

Most nuclear fission reactions are man-made, and the products are often radioactive isotopes; radiation emitted from isotopes produced by man-made nuclear reactions is called 'artificial radiation'.

Nuclear fission is carried out in nuclear reactors to produce nuclear energy. Most nuclear reactors use ${}^{235}U$ which is present in small quantities in naturally occurring uranium. The uranium needs to be 'enriched', which means increasing the amount of ${}^{235}U$, before it is used in nuclear reactors. This is done by bombarding the nucleus with neutrons and break up (i.e. the nuclei undergo nuclear fission). This produces a large amount of heat as well as extra neutrons. The excess heat needs to be cooled, and this is usually done with water.



Figure 9: Koeberg nuclear power plant in Cape Town is cooled by cold water from the Atlantic Ocean pumped through an isolated circuit at 80 tons per second



Figure 10: A cross-section of a nuclear reactor

In a nuclear power station nuclear fuel undergoes a controlled chain reaction in the reactor to produce heat – nuclear energy to heat energy. The chain reaction is controlled by boron control rods. Boron is a non-fissile material. This means it will not undergo a fission reaction when bombarded with neutrons. When the Boron absorbs the neutrons then the chain reaction will slow down due to a lack of neutron-producing reactions.

In times of high energy demand, control rods are raised allowing more reactions to take place. In times of low demand, control rods are lowered to absorb neutrons and hence decrease the number of reactions; less reactions = less heat energy produced.

Heat from the nuclear reaction is used to change water into steam in the boiler. (The water in the boiler NEVER actually comes into contact with the same coolant that is used in the reactor but is transferred from the reactor to the external water by using a heat exchanger.)

The steam drives the turbine (heat energy to kinetic energy - the same as a conventional power station).

This drives the generator to produce electricity – kinetic energy to electrical energy.

- The fuel rods: source of enriched uranium fuel that undergoes fission reactions.
- The moderator: the fuel rods are encased in the moderator (graphite is sometimes used) and help the fission reaction to proceed efficiently.
- The control rods: these are normally made of boron and help control the fission reaction.
- The coolant: this is a liquid or gas that transfers energy from the reactor to the heat exchanger; it's also used to cool the reactor.
- The containment vessel: this is a thick concrete shield that absorbs neutrons and ionising radiation. In practice this can be 5.5 m thick.

There are a number of advantages of nuclear fuels:

- Unlike fossil fuels, nuclear fuels do not produce carbon dioxide or sulphur dioxide.
- In terms of energy produced, 1 kg of nuclear fuel produces the same energy as 2,900,000 kg of coal.
- Since comparatively small volumes of fuel are used, it can be easily transported by road or rail.
- Waste products are small in volume.

There are also disadvantages of nuclear fuels:

- Like fossil fuels, nuclear fuels are non-renewable energy resources.
- If there is an accident, large amounts of radioactive material could be released into the environment. Modern reactor designs are extremely safe, and this very rarely happens.
- Nuclear waste remains radioactive and is hazardous to health for thousands of years. It must be stored safely. Storing nuclear waste is another hotly debated issue. Nuclear waste from Koeberg is stored in the Kalahari Desert.
- An atom bomb is a device which releases a large amount of energy very quickly as a result of a nuclear fission reaction, causing an explosion. Usually plutonium-239 or uranium-235 is used and the reaction is started by firing neutrons at the sample. Because nuclear fission produces radioactive isotopes as products, atom bombs can leave behind radioactive material long after they have exploded.

Nuclear fusion

Nuclear fusion is the joining together of two smaller nuclei to form a single, larger nucleus.

 $\mathsf{E.g.}~^2_1H + {}^3_1H \rightarrow ^4_2He + {}^1_0n$



Figure 11: The three isotopes of hydrogen



Figure 12: Nuclear fusion occurring between two isotopes of hydrogen releases huge amounts of energy

The larger nucleus is usually more stable than the two smaller nuclei, and so a lot of heat energy and gamma rays are released during nuclear fusion. Nuclear fusion takes place naturally in the sun – especially the fusion of hydrogen into helium. Nuclear fusion reactions are the source of the sun's energy, and it is not currently possible to produce energy commercially using nuclear fusion reactions because very high temperatures are needed.

Nuclear fusion does not generally produce radioactive products; as a result it is considered much safer than nuclear fission.

A hydrogen bomb is a device which releases a large amount of energy very quickly because of the nuclear fusion of hydrogen atoms. Due to the large amount of energy required to start nuclear fusion, a nuclear fission reaction is used to create the heat necessary for the nuclear fusion reaction to start.



3. Explain how electricity can be made from nuclear fission and explain why it is not currently possible to make electricity from nuclear fusion.

The <u>full solutions</u> are at the end of the unit.

The use of radioactivity

Although the radiation emitted by radioactive atoms can be dangerous, it also has useful applications, such as for carbon dating and as tracers.

Carbon-dating is done using carbon-14, which is radioactive and has a half-life of 5730 years. Carbon atoms in living organisms are constantly being replaced due to the carbon cycle and so the proportion of carbon existing as carbon- 14 in living organisms is constant. When cells die, the carbon atoms in those cells are no longer replaced and so the amount of carbon-14 in those cells gradually decreases over time. As the carbon-14 decays 5730 years after death, it is possible to estimate how many years ago a sample of tissue died, and this can be used to estimate the age of fossils, skeletons, and some fabrics.

Radioactive material is used in tracers. Because gamma radiation can travel through walls and earth, it is possible to track the location of gamma emitting radioactive material by detecting the gamma rays. Tracers are used in several ways:

- Tracers are used in industry to locate blockages in underground or underwater pipes, a sample of radioactive material is inserted into the pipe and its progress is followed by tracking the radiation it emits. Radiation will accumulate at a blockage and disperse at a leak so engineers can find leaks and blockages without opening up the entire pipe.
- Tracers are used in agriculture to monitor how, and how fast, plants can absorb certain nutrients from the soil. Plants absorb water and other elements from the soil; if radioactive isotopes of these elements are placed in the soil, their rate of absorption can be measured by how much radioactivity is detected in the plant.
- Tracers are used in medicine to monitor the digestive system, respiratory system, or circulatory system; radioactive material can be inhaled, ingested, or injected and its progress through the body can be monitored.
- Cancer cells also absorb certain atoms (for example thyroid cancer absorbs iodine) so radioactive iodine will concentrate in the cancerous area, allowing the cancer to be located and treated. Exposure to large quantities of radiation is known to kill cells or cause them to mutate and become cancerous, but radiation can also be used to kill cancerous cells by using a targeted dose of radiation without harming healthy cells.

Summary

In this unit you have learnt the following:

- $\cdot\,\,$ The major types of radioactivity include alpha particles, beta particles, and gamma rays.
 - Alpha decay is a type of radioactive decay in which a particle with two neutrons and two protons is ejected from the nucleus of a radioactive atom.
 - Beta decay is a type of radioactive decay in which a beta particle (fast energetic electron or positron) is emitted from an atomic nucleus.
 - Gamma decay is a type of radioactive decay that a nucleus can undergo. What separates this type
 of decay process from alpha or beta decay is that no particles are ejected from the nucleus when it
 undergoes this type of decay. Instead, a high energy form of electromagnetic radiation a gamma
 ray photon Is released.
- The half-life of an isotope is used to describe the rate at which the isotope will decay and give off radiation.
- Using the half-life, it is possible to predict the amount of radioactive material that will remain after a given amount of time.
- Fission and fusion are two physical processes that produce massive amounts of energy from atoms.
- Fission occurs when a neutron slams into a larger atom, forcing it to excite and spilt into two smaller atoms. Additional neutrons are also released that can initiate a chain reaction.
 - When each atom splits, a tremendous amount of energy is released.
 - Uranium and plutonium are mostly used for fission reactions in nuclear power reactors because they are easy to initiate and control.
 - The energy released by fission in these reactors' heats water into steam. The steam is used to spin a turbine to produce carbon-free electricity.
- Fusion occurs when two atoms slam together to form a heavier atom, for example, when two hydrogen atoms fuse to form one helium atom.
 - This is the same process that powers the sun and creates huge amounts of energy—several times greater than fission. In addition, it does not produce highly radioactive fission products.
 - Fusion reactions are being studied by scientists but are difficult to sustain for long periods of time because of the tremendous amount of pressure and temperature needed to join the nuclei together.

Unit 1: Assessment

Suggested time to complete: 20 minutes

- 1. Write a balanced nuclear equation to show:
 - a. the emission of an alpha particle from thorium-228
 - b. the emission of a beta particle from oxygen-19.

2.

- a. State, with a reason, which type of radiation is the most ionising and which type is the least ionising.
- b. State, with a reason, which type of radiation is the most penetrating and which type is the least penetrating.
- c. Describe two ways in which exposure to excessive radiation can be harmful to living cells.
- 3. Explain why gamma rays are more penetrating than alpha and beta particles and explain how this makes them more dangerous.
- 4. Write a nuclear equation to show the emission of:
 - a. an alpha particle from bismuth-210
 - b. a beta particle from sodium-24.
- 5. Calculate the time it will take for the activity of a sample of ${}^{131}I$, which has a half-life of 8 days, to form to 12.5% of its original value.
- 6. Carbon-14 has a half-life of 5730 years; estimate the age of a fossil which contains 12.5% of the carbon-14 content of living tissue.
- 7. Uranium-235 can undergo fission to produce caesium-144 and rubidium-90. Work out the number of neutrons released and write a nuclear equation for the reaction.
- 8. Explain the difference between an atom bomb and a hydrogen bomb.
- 9. Give an example of the use of radioactive tracers in:
 - a. agriculture
 - b. industry
 - c. medicine

The <u>full solutions</u> rea at the end of the unit.

Unit: 1 Solutions

Exercise 1.1

1.

- a. two protons and two neutrons; atomic number decreases by ${\mathbf 2}$ and mass number decreases by ${\mathbf 4}$
- b. an electron emitted from the nucleus; atomic number increases by 1
- c. emission of alpha and beta particles reduce potential energy of nucleus; excess energy is emitted as a gamma ray

2. $^{209}_{84}\text{PO} \rightarrow^{4}_{2}\text{He} + ^{205}_{82}\text{Pb}$

Back to Exercise 1.1

Exercise 1.2

- 1. The time taken for the amount or activity of a sample to fall to half of its original value.
- 2. 7.2 mins = 3 half-lives so it will have decayed by 87.5%.

Back to Exercise 1.2

Exercise 1.3

- 1. Breaking up of a nucleus into two or more smaller nuclei.
- 2. Joining of two or more nuclei to make a single nucleus.
- 3. Heat released when nucleus breaks up, this boils water which drives a turbine which drives a generator. Very high temperature are needed to start the fusion reaction.

Back to Exercise 1.3

Unit 1: Assessment

1.

```
a. ^{228}_{90}Th \rightarrow^{224}_{88}Ra +^{4}_{2}He
```

b. ${}^{19}_{8}\text{O} \rightarrow {}^{19}_{9}\text{F} + {}^{0}_{-1}\beta$

2.

- a. Alpha is the most ionising it is large and has a +2 charge. Gamma is the least ionising it has no charge.
- b. Gamma is the most penetrating as it has no mass so does not collide with particles; it has no charge so is not strongly ionising. Alpha is the least penetrating because it is strongly ionising and is destroyed when it ionises another particle.
- c. It can kill/burn cells or cause them to mutate and become cancerous.
- 3. They have no mass or charge so do not interact with other materials; as a result it is very difficult to protect oneself against them.

4.

- a. $^{210}_{83}\text{Bi} \rightarrow ^{206}_{81}\text{Tl} + ^{4}_{2}\text{He}$
- b. $^{24}_{11}$ Na \rightarrow^{24}_{12} Mg + $^{0}_{-1}\beta$
- 5. 12.5% of value = 3 half-lives = 24 days
- 6. 12.5% is three half-lives so age = 5730 x 3 = 17190 years
- 7. ${}^{235}_{92}$ U $\rightarrow {}^{144}_{55}$ Cs + ${}^{90}_{37}$ Rb + ${}^{1}_{0}$ n (1 neutron released)
- 8. Atom bomb energy is released because of nuclear fission. Hydrogen bomb most energy released because of nuclear fusion.

9.

- a. Agriculture: monitoring how fast plants take up certain nutrients.
- b. Industry: identifying blockages and leaks in underground pipes.
- c. Medicine: locating blockages in digestive or circulatory system.

Back to Unit 1: Assessment

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SUBJECT OUTCOME IX MATTER AND MATERIALS: STATE, EVALUATE AND APPLY PROPERTIES OF FLUIDS ON EVERYDAY LIFE AND INDUSTRIAL CONTEXTS



Subject outcome

Subject outcome 5.2: State, evaluate and apply properties of fluids on everyday life and industrial contexts



Learning outcomes

- Define and apply hydrostatic pressure and its relation to depth and density of a liquid.
- Define, give examples of, and use Pascal's principle.
- Describe fluid flow and use a Venturi meter to show the relationship between pressure, crosssection of the pipe changes speed of fluid flow.



Unit 1 outcomes

By the end of this unit you will be able to:

- Define, describe and calculate hydrostatic pressure in relation to the density and depth of a liquid.
- Use Pascal's principle and state common uses of this principle.
- Show the relationship between pressure and the speed of fluid flow using a Venturi meter.

Unit 1: Properties of Fluids

DYLAN BUSA



By the end of this unit you will be able to:

- Define, describe and calculate hydrostatic pressure in relation to the density and depth of a liquid.
- Use Pascal's principle and state common uses of this principle.
- Show the relationship between pressure and the speed of fluid flow using a Venturi meter.

What you should know

Before you start this unit, make sure you can

- Describe a fluid. Refer to level 3 subject outcome 5.4 units 1 and 2 if you need help with this.
- State and explain what density and pressure are. Refer to <u>level 3 subject outcome 5.4 units 1</u> and <u>2</u> if you need help with this.
- Explain and calculate mechanical advantage. Refer to <u>level 3 subject outcome 2.5</u> if you need help with this.
- State and explain the mechanical properties of matter. See <u>level 3 subject outcome 5.1</u> if you need help with this.

Introduction

We know that there are four basic states of matter – solid, liquid, gas and plasma (see figure 1). Together, liquids and gases are called fluids. Fluids take the shape of the container they are in. They can also be made to move from an area of high pressure to an area of low pressure. In other words, fluids can flow.



Figure 1: The four fundamental states of matter – clockwise from top left: solid (ice), liquid (water), lightning (plasma), gas (water vapour)

But we know that liquids and gases have some important differences. For example, while gases take up the whole volume of the container they are in, liquids do not. Gases also tend to be far more compressible. By applying a force, we can greatly decrease the volume that a gas occupies. The volume of a gas also changes far more with respect to temperature than does a liquid's volume. As the temperature of a gas increases, so does its volume.

We focused on many of these properties of gases in <u>level 3 subject outcome 5.4</u>. In this unit, we will focus on liquids and we will make two important assumptions about liquids:

- liquids are **incompressible** their volume cannot be reduced by apply a force
- the **volume** of liquids does not change with temperature.

Now, of course, in the real world, all liquids can be compressed to some degree. It is just that the forces required are usually very, very large. Also, the volume of liquids does increase with a rise in temperature. It is just that any change in the volume is usually very, very small.

Hydrostatic pressure

To understand hydrostatic pressure, we need to make sure that we understand what pressure and density are. Let's start with density.

Density

You may have heard this old trick question: "Which weighs more: a ton of bricks or a ton of feathers?" The answer, of course, is that they both weight exactly the same – a ton $(1 \ 000 \ \text{kg})$ (see figure 2).



The volume of a ton of feathers, however, will be much greater than the volume of a ton of bricks. The 'stuff' of feathers is much less tightly packed than the 'stuff' of bricks. They have different densities.

Density is the mass of a unit volume of a substance. It is a measure of how tightly packed the matter of a substance is. The more tightly packed the matter, the more mass a given volume of the substance will have.

Density (ρ the Greek letter rho) is therefore defined as mass divided by volume or $\rho = \frac{m}{V}$. Density has the SI units of kg.m⁻³.



Every solid, liquid and gas has a specific, defined density at standard temperature and pressure. Table 1 shows the densities of some common substances. Notice that the densities of solids are generally greater than liquids and that the density of liquids is generally greater than gases.

		$\begin{array}{c} \text{Liquids} \\ (\text{at } 0.0 \ C^{\circ}) \end{array}$		Gases (at $0.0~{\rm C^\circ}$ and $101.3~{\rm kPa}$)	
Substance	$ ho~({ m kg.m}^{-3})$	Substance	$ ho~({ m kg.m}^{-3})$	Substance	$ ho~({ m kg.m^{-3}})$
Aluminium	$2.70 imes10^3$	Benzene	$8.79 imes10^2$	Air	$1.29 imes 10^0$
Bone	$1.90 imes10^3$	Blood	$1.05 imes 10^3$	Carbon dioxide	$1.98 imes 10^0$
Brass	$8.44 imes 10^3$	Ethyl alcohol	$8.06 imes 10^2$	Carbon monoxide	$1.25 imes 10^{0}$
Concrete	$2.40 imes 10^3$	Gasoline	$6.80 imes10^2$	Helium	$1.80 imes10^{-1}$
Copper	$8.92 imes10^3$	Mercury	$1.36 imes10^4$	Hydrogen	$9.00 imes 10^{-2}$
Glass	$2.60 imes10^3$	Olive oil	$9.20 imes 10^2$	Methane	$7.2 imes10^{-2}$
Gold	$1.93 imes 10^4$	Water	$1 imes 10^3$	Nitrogen	$1.25 imes 10^{0}$
Granite	$2.70 imes10^3$			Oxygen	$1.43 imes 10^{0}$
Iron	$7.86 imes10^3$				
Lead	$1.13 imes 10^4$				
Oak	$7.10 imes10^2$				
Pine	$2.70 imes 10^3$				
Platinum	$2.14 imes 10^4$				
Polystyrene	$1.00 imes 10^2$				
Uranium	$1.87 imes 10^3$				

Table 1: The densities of some common substances



Example 1.1

A dam has a surface area of 40.0 km^2 and an average depth of 45.0 m. What mass of water is held in the reservoir? Assume the density of water is $1\ 000\ \text{kg.m}^{-3}$.

Solution

The first thing we need to do is convert the surface are of the dam from ${\rm km}^2$ to ${\rm m}^2$.

40.0 km² = 40 × (1 000 × 1 000) = 40 000 000 m² The total volume of water is: V = surface area × depth= 40 000 000 m² × 45.0 m = 1 800 000 000 m³ = 1.8 × 10⁹ m³ $\rho = \frac{m}{V}$ ∴ $m = \rho \times V$ = 1 000 kg.m⁻³ × 1.8 × 10⁹ m³ = 1.8 × 10¹ 2kg

Pressure

All fluids exert a pressure. Pressure is defined as the force divided by the area perpendicular to the force over which the force is applied. In other words, $P = \frac{F}{A}$. Pressure has the SI units of $N.m^{-2}$.

Force is most often measured in pascals (Pa), in recognition of the work Blaise Pascal did in defining and explaining pressure. We define 1 Pa as being equal to $1 N.m^{-2}$.



It is important to note that the forces exerted by fluids because of pressure are only exerted **perpendicular** to surfaces. Fluids cannot exert sideways or shear forces.

Fluid pressure also has no direction. The force can be exerted in any direction. When someone is swimming through water, the pressure is exerted on all sides of the swimmer, since the water would flow into the space they occupy if they were not there. The arrows in figure 3 below represent the directions and sizes of the forces exerted at various points on the swimmer (all perpendicular). Note that the forces are larger

underneath the swimmer, due to greater depth, giving a net upward or buoyant force that is balanced by the weight of the swimmer.



Figure 3: The forces exerted by water on a swimmer in water

Pressure at a depth

We noted above that the forces acting on the underside of the swimmer in figure 3 were greater because the depth was greater. Let's explore the relationship between depth and pressure in more detail.

If you have ever dived into a deep pool or body of water, you might have noticed that, at a certain depth, your ears start to hurt. This is because the deeper you go under water, the greater the pressure exerted by the water. A simple way to think of this is that the deeper you go, the more water there is vertically above you. The more water, the greater the weight of this water, the more force it will be able to exert per unit area.



Therefore, the force with which this column of water pushes down on the bottom of the reservoir will be its weight i.e. $mg = 45 \ 000 \text{kg} \times 9.8 \text{ m.s}^{-2} = 441 \ 000 \text{ N}.$

$$egin{aligned} P &= rac{F}{A} \ &= rac{441\ 000\ \mathrm{N}}{1\ \mathrm{m}^2} &= 441\ 000\ \mathrm{Pa} \ &= 4.41 imes 10^5\ \mathrm{Pa} \end{aligned}$$

We can derive a simple expression for the pressure due to the weight of any fluid. We know from the above example that:

$$P = rac{mg}{A}$$

We can find the mass of the fluid from its volume and density: m=
ho V

The volume of the fluid related to the dimensions of a right prism container where A is the area of the base and h is the vertical height is as follows:

$$V = Ah$$

Therefore, we can say that:

m=
ho AhIf we use this in the expression for pressure, we get: $P=rac{(
ho Ah)g}{}$

$$P = \frac{1}{A}$$

If we cancel the area terms and rearrange some other terms we get:

 $P = h\rho g$



where h is the vertical height of the fluid above the point where the pressure is being measured, ρ is the density of the fluid and g is the acceleration due to gravity (see figure 4).



While you may not realise it, the air around you exerts an enormous pressure on your body. Use the expression above to calculate what this pressure is at sea level if the height of the column of air above you is 8.440 km and the density of air is 1.225 kg.m^{-3} .

Did you get that the air exerts a pressure on your body of about $101 \ 325 \ Pa$ at sea level. Obviously, the higher up you go, the less this **atmospheric pressure** is because there is less air above you. You will sometimes see atmospheric pressure measured in **atmospheres**.

$1 \text{ atmosphere} = 101 \ 325 \ \text{Pa}.$

You don't feel this atmospheric pressure because the pressure inside your body is roughly the same as the atmospheric pressure outside. Therefore, both forces acting on our bodies cancel out and we do not feel uncomfortable.

Now when you are in water, your body does not only experience the pressure due to the water. It also experiences the pressure of the atmosphere pushing down on the surface of the water as well. Do the following activity to learn more.



1. Open the interactive simulation called <u>Under Pressure</u>



You will find four icons listed in the bottom left-hand corner. Make sure the top icon is selected. Make sure also that the Fluid Density is set to 1 000 kg.m⁻³ (in other words so that we are dealing with water) and the Gravity is set to 'Earth'.

- 2. Click and drag the pressure gauge to measure the air pressure directly above the grass and above the surface of the water. What is the pressure in each case? Why are they different readings?
- 3. Now click on the Grid checkbox to turn the grid on. You should see that the level of the water is about 1.5 m. Calculate the pressure at the bottom of the tank if this pressure is the combination of the atmospheric pressure at the surface of the water and the pressure due to the weight of the water itself. Use your pressure gauge to confirm.
- 4. Click on the top stopper to fill the tank with water. Recalculate the pressure at the bottom of the tank due to the atmospheric pressure and the pressure due to the weight of the water. Use your pressure gauge to confirm.
- 5. Now click on the third icon on the left-hand side. Make sure once again that the density and gravity are $1\ 000\ {\rm kg.m^{-3}}$ and 'Earth' respectively.
- 6. First, measure the pressure at the bottom of both parts of the tank. Is the pressure the same. Why?
- 7. Place the pressure gauge at the bottom of the left-hand part of the tank. You will see some masses on the left-hand side. Drag the 500 kg mass and drop it into the dotted box as indicated. What happens to the pressure at the bottom of the left-hand part of the tank? Why is this? What happens to the water level in the right-hand part of the tank? Why is this? Are the pressures at the bottom of both halves of the tank still the same? Why?
- 8. What happens when you add either or both of the other masses?

What did you find?

1. This is what your screen should look like.



- 2. The pressure just above the grass is 101.325 kPa. The pressure just above the surface of the water is 101.524 kPa. The pressure at the surface of the water is greater than just above the ground because the surface of the water is below the ground and so the column of air exerting this force is higher.
- 3. Pressure at surface of the water is 101.524 kPa. The pressure at the bottom of the tank due to the weight of the water is calculated as follows:

$$P=h
ho g$$

 $= 1.5 \ {\rm m} \times 1 \ 000 \ {\rm kg.m^{-3}} \times 9.8 \ {\rm m.s^{-2}} \qquad = 14 \ 700 \ {\rm Pa}$

$$= 14.7$$
 kPa

Therefore, the total pressure at the bottom of the tank is 101.524 kPa + 14.7 kPa = 116.224 kPa.

The pressure gauge reads 115.968 kPa. While it is not exact, it is very close.

4. If we fill the tank, the depth of the water will be 3 m. The pressure at the surface of the water will be 101.325 kPa. The pressure at the bottom of the tank due to the weight of the water will be: $P = h\rho g$

 $= 3 \ m \times 1 \ 000 \ kg.m^{-3} \times 9.8 \ m.s^{-2} \qquad = 29 \ 400 \ Pa$

= 29.4 kPa

Therefore, the total pressure at the bottom of the tank is 101.355 kPa + 29.4 kPa = 130.725 kPa.

The pressure gauge reads 130.650 kPa. While it is not exact, it is very close.

5. This is what your screen should look like.



- 6. The pressure at the bottom of both parts of the tank is 123.798 kPa. It is the same because the height of the water vertically above each bottom is exactly the same. Therefore, the atmospheric pressure at the surface and the pressure at the bottom due to the weight of the water is the same.
- 7. The mass pushes down the surface of the water. Therefore, the pressure at the surface is now greater. It is atmospheric pressure plus the weight of the 500 kg mass. Even though the height of the water above the bottom of the tank is lower, the pressure still increases because the weight of the mass is greater than the volume of water being displaced. The solid mass has a density greater than water. The water level in the right-hand part of the tank rises. The water being displaced on the left has to go somewhere. The water is non-compressible. The pressure at the bottom of the left-hand part of the tank. This pressure increases because the height of the water has increased and the weight of the water has increased by a greater amount than the decrease in the atmospheric pressure at the surface.
- 8. As you place additional masses, the pressure in the tank increases and the water level on the righthand side increases.

In activity 1.1, we saw that the pressure at the bottom of a container of fluid exposed to the atmosphere is the sum of the pressure due to the weight of the fluid and the atmospheric pressure at the surface of the fluid.

It is important to note that the shape of the container has no effect on the pressure. All that matters is the density of the fluid and its vertical height. Have a look at each of the pressure gauges in figure 5. In all three cases the shape of the container is different but the level of the liquid is the same and the pressure gauges all show the same reading. This is because pressure is the force per unit area acting perpendicular to a surface. Therefore, we are only concerned with the height of a vertical column of fluid above the surface.



Figure 5: Pressure depends on depth not shape

We also learnt that when a force is exerted on a fluid, this changes the pressure the fluid exerts and that this change in pressure is transmitted throughout the fluid. This is an important principle we will turn our attention to next.



- d. the distance below the surface
- e. all of the above
- 4. In the equation for pressure P=
 ho imes g imes h, the units for g (SI system) are:

- a. $kg.m^{-3}$
- b. $\mathrm{m.s}^{-1}$
- C. $kg.m.s^{-1}$
- d. $m.s^{-2}$
- 5. What is the pressure at the bottom of a swimming pool that is 3 metres deep?
 - a. $(101 imes 10^5) + (2.94 imes 10^4)$ Pa
 - b. $(101 \times 10^5) + (1.09 \times 10^5)$ Pa
 - c. $(101 \times 10^5) + (3.63 \times 10^4)$ Pa
 - d. $(101 \times 10^5) + (7 \times 10^4)$ Pa
- 6. A pressure gauge is placed at the bottom of a reservoir of unknown depth. The reservoir contains water. If the pressure registered on the gauge is 126.578 kPa and atmospheric pressure is 101.303 kPa, how deep is the reservoir?
- 7. An open container of an unknown liquid is at sea level. A pressure gauge at a depth of 1.5 m measures a pressure of 126.205 kPa. What is the fluid density?

The <u>full solutions</u> are at the end of the unit.

Pascal's principle

We ended the section above by noting that when a force is exerted on a fluid, this changes the pressure the fluid exerts and that this change in pressure is transmitted throughout the fluid. This is a simple statement of Pascal's principle. This principle is the reason why a hydraulic jack can lift a car (see figure 6).



But how can a fluid be used to lift such heavy objects?

We know that pressure is defined as force per unit area. We also know that the pressure in a fluid can be increased by pushing directly on the fluid. This is how the heart pumps blood around our bodies. The key though is to have a closed system. If you try to push on a fluid in an open system, like a river, the fluid just flows away.

In an enclosed system, the fluid cannot flow away, and so the applied force can more easily result in an increase in pressure. When such a force is applied, the atoms in the fluid transmit the pressure to all parts of the fluid and to the walls of the container. Importantly, this change in pressure is transmitted undiminished (without any decrease).

This phenomenon is called Pascal's principle. It states that a change in pressure applied to an enclosed fluid is transmitted undiminished to all portions of the fluid and to the walls of its container.



But we still don't know how this change in pressure results in the ability to lift a car? Consider the system in figure 7. Here we have a closed system of two connected containers each with a stopper. Because they are connected, any change in pressure in one container will be transmitted undiminished through the whole system.



Figure 7: Pascal's principle used to exert a force

We know, therefore, that at all times $P_1 = P_2$. But we also know that:

$$P_1=rac{F_1}{A_1}$$
 and $P_2=rac{F_2}{A_2}$

Therefore:

 $\frac{F_1}{A_1} = \frac{F_2}{A_2}$ $\therefore \frac{F_1}{F_2} = \frac{A_1}{A_2}$

Now if $A_2 = 2A_1$, this means that $F_2 = 2F_1$. In other words, we will have doubled the force exerted by the liquid on the right-hand plunger in terms of the force we exerted on the left-hand plunger. We have a force multiplier. This simple hydraulic system has a mechanical advantage of two.

Note

Spend some time playing with the simple hydraulic system simulation called <u>Pascal's Principle in a</u> <u>Hydraulic Lift</u>. Here you can change the radius of each plunger as well as the output force.





Example 1.4

In a car's hydraulic brake system, a force of 500 N is applied via the brake pedal, which acts on the master cylinder. The change in pressure created in the master cylinder is transmitted to each of the four slave cylinders at the wheels. If the master cylinder has a diameter of 0.500 cm, and each slave cylinder has a diameter of 2.50 cm, calculate the force created at each of the slave cylinders.

Solution

In a hydraulic system we know that $rac{F_1}{F_2}=rac{A_1}{A_2}.$ We need to first calculate the areas of the cylinders.

Master cylinder: $A = \pi r^2 = \pi imes (0.25 \ {
m cm})^2 = 0.0625 \ {
m cm}^2$

Slave cylinders: $A=\pi r^2=\pi imes (1.25~{
m cm})^2=1.5625~{
m cm}^2$

$$egin{array}{ll} rac{F_1}{F_2} &= rac{A_1}{A_2} \ dots &: F_2 &= rac{A_2}{A_1} imes F_1 \ &= rac{1.5625 \ \mathrm{cm}^2}{0.0625 \ \mathrm{cm}^2} imes 500 \ \mathrm{N} \ &= 1.25 imes 10^4 \ \mathrm{N} \end{array}$$

Alternatively, we could have done the calculation as follows:

$$egin{aligned} rac{F_1}{F_2} &= rac{A_1}{A_2} \ dots, F_2 &= rac{A_2}{A_1} imes F_1 \ &= rac{\pi r_2^2}{\pi r_1^2} imes F_1 \ &= rac{(2.5 ext{ cm})^2}{(0.500 ext{ cm})^2} imes 500 ext{ N} \ &= 1.25 imes 10^4 ext{ N} \end{aligned}$$

Did you know?

Hydraulic systems are used for many applications. They are used in lift and crane systems, in the braking systems of motor vehicles, to operate the landing gear and wing flaps of aircraft and in many industrial robots.

Watch the video called Pascal Law | Hydraulic System and Pascal law | applications of pascal's law.

Pascal Law | Hydraulic System and Pascal law | applications of pascal's law (Duration: 03.41)





Exercise 1.2

- 1. What force must be exerted on the master cylinder of a hydraulic lift to support the weight of a 2 500 kg car resting on a secondary cylinder? The master cylinder has a 2.00 cm diameter and the secondary cylinder has a 25.0 cm diameter.
- 2. A certain hydraulic system is designed to exert a force 120 times as large as the one put into it.
 - a. What must be the ratio of the area of the cylinder that is being controlled to the area of the master cylinder?
 - b. What must be the ratio of their diameters?
 - c. By what factor is the distance through which the output force moves reduced relative to the distance through which the input force moves? Assume no losses due to friction.

The <u>full solutions</u> are at the end of the unit.

Fluid pressure and velocity

So far in this unit, we have considered the pressure exerted by a fluid that is not moving. But fluids that are moving can also be under pressure and exert pressure. Indeed, it is differences in pressure that make a fluid move through a pipe in the first place – from high pressure to lower pressure.

However, it turns out that the thickness of the pipe affects both the pressure of the fluid and the velocity with which it flows in some surprising ways. Consider the case shown in figure 8.



Figure 8: A tube with a narrowing

The first thing we need to note is that the same volume of fluid needs to flow through each part of the tube each second. Let's assume that the shaded volume of fluid in the wide part of the tube flows past **point 1** in one second. When this volume of fluid gets to the narrow part of the pipe, the same volume of fluid will have to flow past **point 2** also in one second. However, we can see that, because the diameter of the tube has decreased, the length of this column of fluid needs to increase. This means that the velocity of the fluid needs to increase in order to get this 'longer' column of fluid past point 2 in one second.

So, as the diameter of a tube **narrows**, the velocity of the fluid flowing through that part of the tube **increases**.

We call the volume of fluid flowing through a tube per unit time the **flow rate** (*Q*). The SI units for flow rate are $m^3 \cdot s^{-1}$ but it is also commonly measured in $L \cdot min^{-1}$. The flow rate can be calculated as the cross-sectional area of the tube multiplied by the velocity of the fluid. In other words, Q = Av (see figure 9).



Figure 9: Flow rate through a tube

Look back at figure 8. We know that the flow rate at point 1 must be equal to the flow rate at point 2. Therefore, we can say that if $Q_1 = Q_2$ then $A_1v_1 = A_2v_2$.

This is called the **equation of continuity** and is valid for any incompressible fluid with constant density. This can be seen when water flows from a hose. The smaller the diameter of the nozzle, the greater its velocity.



Example 1.5

A nozzle with a diameter of 0.5 cm is attached to a high-pressure hose with a radius of 2 cm. The flow rate through hose and nozzle is $0.6 L.s^{-1}$. Calculate the speed of the water:

- 1. in the hose.
- 2. in the nozzle.

Solutions

1. We can use the concept of flow rate to calculate the velocity of the water through the hose.

$$egin{aligned} Q &= Av \ & \ddots v &= rac{Q}{A} \ & = rac{0.6 \,\, \mathrm{L.s^{-1}} imes 10^{-3} \,\, \mathrm{m^3.L^{-1}}}{\pi \, (0.01 \,\, \mathrm{m^2})} \ & = 1.91 \,\, \mathrm{m.s^{-1}} \end{aligned}$$

2. We can use the equation of continuity to calculate the velocity of the water flowing through the nozzle.

$$egin{aligned} A_1 v_1 &= A_2 v_2 \ dots v_2 &= rac{A_1}{A_2} imes v_1 \ &= rac{\pi r_1^2}{\pi r_2^2} imes v_1 \ &= rac{r_1^2}{\pi r_2^2} imes v_1 \ &= rac{r_1^2}{r_2^2} imes v_1 \ &= rac{(2 \ \mathrm{cm})^2}{(0.25 \ \mathrm{cm})^2} imes 1.91 \ \mathrm{m.s^{-1}} \ &= 122.24 \ \mathrm{m.s^{-1}} \end{aligned}$$

Now, what do you think happens to the pressure of the fluid as it flows through the narrow part of the tube in figure 8? Does its pressure increase or decrease? Spend a few moments thinking about this.

Watch the demonstration called Bernoulli's Principle Demo: Venturi Tube so see what happens to the pressure of a fluid as it flows through a narrower part of a tube. The result may surprise you.



Did you see that as the velocity of the fluid increases through the narrower part of the tube, its pressure **decreased**? This is a simple statement of **Bernoulli's principle** and is the fundamental reason that planes are able to fly.


Bernoulli's principle states that an increase in the speed of a fluid occurs simultaneously with a decrease in pressure or a decrease in the fluid's potential energy.

Did you know?

Bernoulli's principle is named after Daniel Bernoulli who published it in 1738. Leonhard Euler was the first to describe the principle in a mathematical equation in 1752. The Bernoulli equation is

$$P_1+rac{1}{2}
ho v_1^2+
ho gh_1=P_2+rac{1}{2}
ho v_2^2+
ho gh_2$$

where:

 P_1 and P_2 are the pressures at points 1 and 2 v_1 and v_2 are the velocities of the fluid at points 1 and 2 h_1 and h_2 is the height of points 1 and 2 ρ is the density of the fluid

Notice how similar the $\frac{1}{2}\rho v^2$ parts of the expression are to that for kinetic energy $KE = \frac{1}{2}mv^2$. Isn't science amazing!

Watch the video called Venturi effect and Pitot tubes to see how we can use Bernoulli's equation to prove that the greater the velocity of a fluid, the lower its pressure.

Venturi effect and Pitot tubes (Duration: 09.57)



Exercise 1.3

- 1. Calculate the time it would take to fill a public swimming pool with a capacity of 80 000 L using a garden hose delivering 60 L.min^{-1} .
- 2. Water is moving at a velocity of $2 m s^{-1}$ through a hose with an internal diameter of 1.60 cm.
 - a. What is the flow rate in litres per second?
 - b. The fluid velocity in this hose's nozzle is $_{15}~\mathrm{m.s^{-1}}$. What is the nozzle's inside diameter?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Density is mass per unit volume: $\rho = \frac{m}{V}$
- The SI unit of density is $kg.m^{-3}$.
- \cdot Generally, solids are more dense than liquids and liquids are more dense than gases.
- Pressure is force divided by the perpendicular unit area: $P = \frac{F}{A}$
- · The SI unit of pressure is Pa.
- The pressure due to the weight of a fluid is given by $P = h\rho g$ where h is the vertical height of the fluid above the point where the pressure is being measured, ρ is the density of the fluid and g is the acceleration due to gravity.
- The pressure at the bottom of a container of fluid exposed to the atmosphere is the sum of the pressure due to the weight of the fluid and the atmospheric pressure at the surface of the fluid.
- Pascal's principle states that a change in pressure applied to an enclosed fluid is transmitted undiminished to all portions of the fluid and to the walls of its container.
- According to Pascal's principle $rac{F_1}{F_2}=rac{A_1}{A_2}.$
- Flow rate (Q =) is the volume of a fluid moving past a certain point per unit time. The SI units for flow rate are $m^3.s^{-1}$ but it is also commonly measured in $L.min^{-1}$.
- The flow rate can be calculated as the cross-sectional area of the tube multiplied by the velocity of the fluid: Q = Av
- Bernoulli's principle states that an increase in the speed of a fluid occurs simultaneously with a decrease in pressure or a decrease in the fluid's potential energy.
- The relationship between fluid velocity and fluid pressure can be shown with a Venturi meter.

Unit 1: Assessment

Suggested time to complete: 45 minutes

Question 1 taken from Physical Science level 4 paper 2 March 2019 question 4

1. A hydraulic press consists of two plungers A and B in contact with a liquid. The surface area of plunger A in contact with the liquid is 0.1 m^2 and the surface area of plunger B in contact with the liquid is 0.35 m^2 , as shown in the sketch below. A force of 20 N is applied to plunger A in order to hold a block of unknown mass stationary on plunger B.



- a. Calculate the pressure exerted by plunger A on the surface of the liquid.
- b. Is the pressure exerted by the liquid on plunger B, greater than, equal to or less than the pressure exerted by plunger A on the liquid? Explain the answer with reference to a relevant scientific principle.
- c. Calculate the mass of block X.
- 2. A host pours the remnants of several bottles of wine into a jug after a party and inserts a cork with a 2.00 cm diameter into the bottle, placing it in direct contact with the wine. When the host hits the cork into place, the bottom of the jug (diameter 15 cm) breaks away. Calculate the extra force exerted against the bottom if he hit the cork with a 100 N force.

Question 3 taken from Physical Science level 4 paper 2 March 2019 question 4

3. Water from a pipe of varying cross-sectional area flows into a circular steel tank. The cross-sectional area of the pipe at A is 0.08 m^2 and the speed of the water is 2 m.s^{-1} . At B the cross-sectional area is 0.02 m^2 . The water completely fills the pipe at all times.



- a. Calculate the magnitude of the velocity of the water at B.
- b. What volume of water is ejected at B in 30 minutes?
- c. Determine the height of the water in the tank in 30 minutes.
- d. List two applications of Pascal's principle.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

- 1. The equation in option b is incorrect: $ho = rac{m}{V}$
- 2. Option c is correct.
- 3. Option d is correct: $P = \rho gh$. The density of the fluid and the acceleration due to gravity are constant. Therefore, the pressure depends only on the depth h or the distance below the surface.
- 4. Option d is correct. The SI units for acceleration are ${\rm m.s^{-2}}.$
- 5. Option a is correct: $P_{water} =
 ho gh = 1~000~{
 m kg.m^{-3}} imes 9.8~{
 m m.s^{-2}} imes 3~{
 m m} = 2.94 imes 10^4~{
 m Pa}$
- 6. The pressure due to the weight of the water is:
 - 126.578 kPa 101.303 kPa = 25.275 kPa = 25.275 Pa

$$P = \rho g h$$

$$egin{array}{lll} \therefore h = rac{1}{
ho g} \ = rac{25 \; 275 \; {
m Pa}}{1 \; 000 \; {
m kg.m}^3 imes 9.8 \; {
m m.s}^{-2}} \ = 2.58 \; {
m m} \end{array}$$

7. Atmospheric pressure at sea level is 101.325 kPa. Therefore, the pressure due to the weight of the fluid is:

126.205 kPa − 101.325 kPa = 2 488 Pa $P = \rho gh$ $\therefore \rho = \frac{P}{gh}$ $= \frac{2 488 \text{ Pa}}{times9.8 \text{ m.s}^{-2} × 1.5 \text{ m}}$ $= 169.25 \text{ kg.m}^{-3}$

Back to Exercise 1.1

Exercise 1.2

1.

$$egin{aligned} rac{F_1}{F_2} &= rac{A_1}{A_2} \ dots &: F_1 &= rac{A_1}{A_2} F_2 \ &= rac{\pi (1 \ ext{cm})^2}{\pi (12.5 \ ext{cm})^2} imes 24 \ 500 \ ext{N} \ &= 156.8 \ ext{N} \end{aligned}$$

2.

- a. If the ratio of input to output force is 1:120 then the ratio of the area of the input cylinder to the output cylinder must also be 1:120.
- b. The ratio of input cylinder diameter to the output cylinder diameter must be $1:\sqrt{120}$.
- c. $F_1 \cdot d_1 = F_2 \cdot d_2$. If the ratio $F_1 : F_2 = 1 : 120$, then the ratio $d_1 : d_2 = 120 : 1$.

Back to Exercise 1.2

Exercise 1.3

1. $\frac{80\ 000\ \text{L}}{60\ \text{L.min}^{-1}} = 1\ 333.33\ \text{min} = 22\ \text{hrs and} 13\ \text{min.}$ 2. a. Q = Av $= \pi (0.8\ \text{m})^2 \times 2\ \text{m.s}^{-1}$ $= 4.021\ \text{m}^3.\text{s}^{-1}$ $= 4\ 021\ \text{L.s}^{-1}$ b.

$$A_{1}v_{1} = A_{2}v_{2}$$

$$\therefore \frac{A_{1}}{A_{2}} = \frac{v_{2}}{v_{1}} = \frac{15 \text{ m.s}^{-1}}{2 \text{ m.s}^{-1}} = 7.5$$

$$\therefore \frac{\pi r_{1}^{2}}{\pi r_{2}^{2}} = 7.5$$

$$\therefore \frac{r_{1}^{2}}{r_{2}^{2}} = 7.5$$

$$\therefore r_{2}^{2} = \frac{r_{1}^{2}}{7.5}$$

$$= \frac{(0.8 \text{ m})^{2}}{7.5}$$

$$\therefore r_{2} = \sqrt{\frac{(0.8 \text{ m})^{2}}{7.5}}$$

$$= 0.29 \text{ m}$$

Back to Exercise 1.3

Unit 1: Assessment

1.

a.
$$P = rac{F}{A} = rac{20 \ \mathrm{N}}{0.1 \ \mathrm{m}^2} = 200 \ \mathrm{Pa}$$

b. The pressure is the same. Pascal's principle states that a change in pressure applied to an enclosed fluid is transmitted undiminished to all portions of the fluid and to the walls of its container.

C.

$$\frac{F_1}{F_2} = \frac{A_1}{A_2}$$

$$\therefore F_2 = \frac{A_2}{A_1}F_1$$

$$= \frac{0.35 \text{ m}^2}{0.1 \text{ m}^2} \times 20 \text{ N}$$

$$= 70 \text{ N}$$

$$F = mg$$

$$\therefore m = \frac{F}{g}$$

$$= \frac{70 \text{ N}}{9.8 \text{ m.s}^{-2}}$$

$$= 7.14 \text{ kg}$$

$$\frac{F_1}{F_2} = \frac{A_1}{A_2}$$

$$\therefore F_2 = \frac{A_2}{A_1}F_1$$

$$= \frac{\pi (0.075 \text{ m})^2}{\pi (0.01 \text{ m})^2} \times 100 \text{ N}$$

$$= 5 625 \text{ N}$$

3.

2.

a.

$$egin{aligned} &A_1 v_1 &= A_2 v_2 \ dots &v_2 &= rac{A_1}{A_2} v_1 \ &= rac{0.08 \ \mathrm{m}^2}{0.02 \ \mathrm{m}^2} imes 2 \ \mathrm{m.s}^{-1} \ &= 8 \ \mathrm{m.s}^{-1} \end{aligned}$$

$$egin{aligned} Q &= Av \ &= 0.02 \,\, \mathrm{m}^2 imes 8 \,\, \mathrm{m.s^{-1}} \ &= 0.16 \,\, \mathrm{m}^3.\mathrm{s}^{-1} \ 0.16 \,\, \mathrm{m}^3.\mathrm{s}^{-1} imes (30 imes 60) \,\, \mathrm{s} &= 288 \,\, \mathrm{m}^3 \ &= 288 \,\, 000 \,\, \mathrm{L} \end{aligned}$$

c.

b.

$$V = A imes h$$

 $\therefore h = rac{V}{A}$
 $= rac{288 ext{ m}^3}{30 ext{ m}^2}$
 $= 9.6 ext{ m}$

d. Hydraulic lifts and brakes systems for motor vehicles

Back to Unit 1: Assessment

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SUBJECT OUTCOME X MATTER AND MATERIALS: IDENTIFY AND APPLY KNOWLEDGE OF ORGANIC MOLECULES IN EVERYDAY LIFE AND INDUSTRIAL CONTEXT



Subject outcome

Subject outcome 5.3: Identify and apply knowledge of organic molecules in everyday life and industrial context



Learning outcomes

- Define hydrocarbon and organic molecules and identify their use as fuels.
- Identify fractional distillation of crude oil as a source of different fuels.
- Identify physical property as the effect of chain length and give examples.
- Identify functional groups, saturated and unsaturated structures of organic molecules and relate to chemical properties.
- Draw structural formulae and name (IUPAC) organic molecules.
- Describe plastics and polymers refer to related South African industry.
- Describe thermoplastics and thermosets.



Unit 1 outcomes

By the end of this unit you will be able to:

Identify organic molecular structures with functional groups, saturated and unsaturated structures.



Unit 2 outcomes

By the end of this unit you will be able to:

- Understand isomers.
- Name (IUPAC) organic molecules and write their structural formulae.



Unit 3 outcomes

By the end of this unit you will be able to:

• Understand physical and chemical properties of organic molecules.



Unit 4 outcomes

By the end of this unit you will be able to:

- Understand that crude oil is distilled into different components.
- Understand the process of fractional distillation and the uses of the different compounds separated from crude oil.



Unit 5 outcomes

By the end of this unit you will be able to:

- Understand how plastics and polymers are made and refer to related South African industry.
- Understand the difference between thermoplastics and thermosets.

Unit 1: Organic Molecules

EMMA HARRAGE



Introduction

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In this unit you will learn about organic molecules. Organic molecules contain carbon and hydrogen. They usually form long chain structures which are fuels, alcohols, and plastics. Most organic molecules used are distilled from crude oil.

What are organic molecules?

Organic chemistry is the branch of chemistry that deals with organic molecules. An organic molecule is one which contains carbon, although not all compounds that contain carbon are organic molecules. For example, pure carbon compounds such as diamond and graphite are not organic compounds.

Organic molecules can range in size from simple molecules to complex structures containing thousands of atoms! Although carbon is present in all organic compounds, other elements such as hydrogen (H), oxygen (O), nitrogen (N), sulfur (S) and phosphorus (P) are also common in these molecules.

Organic compounds are very important in daily life, and they range from being simple to extremely complex. Organic molecules make up a big part of our own bodies, they are in the food we eat and in the clothes we wear. Organic compounds are also used to make products such as medicines, plastics, washing powders, dyes, along with an extensive list of other items. There are millions of organic compounds found in nature, as well as millions of synthetic manufactured organic compounds.

Organic structure

Carbon has a number of unique properties which influence how it behaves and how it bonds with other atoms:



means that each carbon atom can form a maximum of four bonds with other atoms

Because of the number of bonds that carbon can form with other atoms, organic compounds can be very complex. Carbon can form bonds with other carbon atoms to form single, double, or triple covalent bonds. Carbon can also form bonds with other atoms such has hydrogen, oxygen, nitrogen, and the halogens.

Carbon can bond to form straight chain, branched, and cyclic molecules. Because of this, long chain structures can form. This is known as catenation – the bonding of atoms of the same element into longer chains. These chains can either be unbranched (see figure 2) or branched (see figure 3) and can contain single carbon-carbon bonds only, or double and triple carbon-carbon bonds as well.



Figure 2: Unbranched carbon chains with (a) single carbon-carbon bonds, (b) single and double carbon-carbon bonds and (c) single and triple carbon-carbon bonds



Figure 3: Branched carbon chains, circled in purple, with (a) single carbon-carbon bonds, (b) single and double carbon-carbon bonds and (c) single and triple carbon-carbon bonds

Most of the bonds that carbon forms with other atoms are covalent. Think, for example, of a C–C bond. The difference in electronegativity between the two atoms is zero, so this is a pure covalent bond. In the case of a C–H bond, the difference in electronegativity between carbon (2, 5) and hydrogen (2, 2) is so small that C–H bonds are almost purely covalent. The result of this is that most organic compounds are non-polar. This affects some of the properties of organic compounds.

The formulae of organic compounds

There are five ways of representing organic compounds:

• Structural formula: The structural formula of an organic compound shows every bond between every atom in the molecule. Each bond is represented by a line. The structural formulae of 2-methylpropane and butane.



Figure 4: The different structures of C_4H_{10}



Figure 5: Molecules are not two-dimensional, here are a few examples of different ways to represent methane ($CH_4)$ and ethane ($C_2H_6)$

• Semi-structural formula: It is possible to understand the structure of an organic molecule without writing out all the carbon-hydrogen bonds.



Figure 6: This way of writing a structure is called a semi-structural formula; (a) is 2-methylpropane and (b) is butane

 Condensed structural formula: It is also possible to represent a molecule without showing any bonds between atoms at all. As for a semi-structural representation, the carbon atoms are grouped with the hydrogen atoms bonded directly to each one. The bonds between these groups are not shown. Branched or substituent groups are shown in brackets after the carbon atom to which they are bonded. $CH_3CH(CH_3)CH_3$ – 2-methylpropane (the brackets around (CH_3) represent the branched section of the structural formula

 $CH_3CH_2CH_2CH_3$ – butane Butane can also be written in a more condensed way: $CH_3(CH_2)_2CH_3$

• Molecular formula: This formula shows the actual number of each type of atom. The number of each atom is written as a subscript after the atomic symbol. The molecular formula of 2-methylpropane is: C_4H_{10} . This means that each molecule of 2-methylpropane consists of four carbon atoms and ten hydrogen atoms. The molecular formula of butane is also C_4H_{10} . The molecular formula gives no structural information about the compound.



Functional groups

The way in which a compound will react is determined by a particular characteristic of a group of atoms and the way they are bonded (e.g. double C–C bond, C–OH group). This is called the functional group. This group is important in determining how a compound will react. The same functional group will undergo the same or similar chemical reaction(s) regardless of the size of the molecule it is a part of. Molecules can have more than one functional group.

In one group of organic compounds, called the hydrocarbons, the single, double, and triple bonds between carbon atoms give rise to the alkanes, alkenes, and alkynes, respectively. The double carbon-carbon bonds (in the alkenes) and triple carbon-carbon bonds (in the alkynes) are examples of functional groups.

In another group of organic compounds, called the alcohols, an oxygen and a hydrogen atom are bonded to each other to form the functional group (in other words an alcohol has an OH in it). All alcohols will contain an oxygen and a hydrogen atom bonded together and attached to the carbon chain.

Name of group	Functional group	Example	Structural Formula
Alkane	 	Eth <i>ane</i>	Н Н Н С С Н Н Н
Alkene	c=c	Eth <i>ene</i>	
Alkyne	—c≡c—	Eth <i>yne</i>	н—с≡с—н
Haloalkane/alkyl halide	— <mark>c — X</mark> (X=F,Cl,Br,I)	Chloromethane	H HC H
Alcoh <i>ol /</i> alkan <i>ol</i>	 	Methan <i>ol</i>	<mark>0—Н</mark> н—С—Н Н

Figure 7: Major classes of organic compounds

There are some important points to note as we discuss functional groups:

The beginning of a compound name (prefix) comes from the number of carbons in the longest chain:

Prefix	Number of carbon atoms
meth-	1
eth-	2
prop-	3
but-	4

The end of a compound name (suffix) comes from the functional group, e.g. an alkane has the suffix -ane.



Saturated and unsaturated hydrocarbons

Hydrocarbons that contain only single bonds are saturated compounds and are called saturated hydrocarbons because each carbon atom is bonded to as many hydrogen atoms as possible. Figure 8 shows a molecule of ethane, which is a saturated hydrocarbon.

Hydrocarbons that contain double or triple bonds are unsaturated compounds and are called unsaturated hydrocarbons because they do not contain the maximum number of hydrogen atoms.



Figure 8 shows molecules of ethane and ethene. If you compare the number of carbon and hydrogen atoms in a molecule of ethane and a molecule of ethene, you will see that the number of hydrogen atoms in ethene is less than the number of hydrogen atoms in ethane even though they both contain two carbon atoms. For an unsaturated hydrocarbon compound to become saturated, one of the two (or three) bonds in a double (or triple) bond must be broken, and additional atoms added.

Hydrocarbons

The hydrocarbons that we are going to look at are called aliphatic compounds. The aliphatic compounds are divided into acyclic compounds (chain structures) and cyclic compounds (ring structures). The chain structures are further divided into structures that contain only single bonds (alkanes), those that contain at least one double bond (alkenes) and those that contain at least one triple bond (alkynes).

Unit 1: Organic Molecules | 333

Cyclic compounds (which will not be covered in this book) include structures such as a cyclopentane ring, which is found in insulating foam and in appliances such as fridges and freezers.



Figure 9: The classification of the hydrocarbons

The alkanes

The alkanes are the simplest organic compounds, which are composed of carbon and hydrogen atoms only. Hydrocarbons with only carbon-to-carbon single bonds (C–C), and existing as a continuous chain of carbon atoms bonded to hydrogen atoms only, are called alkanes (or saturated hydrocarbons).



Methane (CH_4), ethane (C_2H_6), and propane (C_3H_8) are the beginning of a series of compounds in which any two members in a sequence differ by one carbon atom and two hydrogen atoms—namely, a CH_2 unit. Any family of compounds in which adjacent members differ from each other by a definite factor (here a CH_2 group) is called a homologous series. The members of such a series have properties that vary in a regular and predictable manner. The principle of homology gives organisation to organic chemistry in much the same way that the periodic table gives organisation to inorganic chemistry.



Table 1: The first seven members of the alkane series

Name	Molecular formula ($\mathrm{C_nH_{2n}}{+2}$)	Condensed structural formula
Methane	CH_4	$ m CH_4$
Ethane	C_2H_6	$\mathrm{CH}_3\mathrm{CH}_3$
Propane	C_3H_8	$\mathrm{CH}_3\mathrm{CH}_2\mathrm{CH}_3$
Butane	C_4H_{10}	$\mathrm{CH}_3\mathrm{CH}_2\mathrm{CH}_2\mathrm{CH}_3$
Pentane	C_5H_{12}	$\mathrm{CH}_3\mathrm{CH}_2\mathrm{CH}_2\mathrm{CH}_2\mathrm{CH}_3$
Hexane	$ m C_6H_{14}$	$\rm CH_3 CH_2 CH_2 CH_2 CH_2 CH_3$
Heptane	C_7H_{16}	$\mathrm{CH}_{3}\mathrm{CH}_{2}\mathrm{CH}_{2}\mathrm{CH}_{2}\mathrm{CH}_{2}\mathrm{CH}_{2}\mathrm{CH}_{3}$

The general formula for alkanes is $C_nH_{2\,n}+2$. Using this formula, we can write a molecular formula for any alkane with a given number of carbon atoms. For example, an alkane with eight carbon atoms has the molecular formula $C_8H_{(2\times8)}$ $_{+\ 2}$ = C_8H_{18} which is called octane.

The alkanes are the most important source of fuel in the world and are used extensively in the chemical industry. Alkanes that contain four or less carbon atoms are gases (e.g. methane and ethane). Longer chains are liquid fuels e.g. octane, which is a key component of petrol and very long chains are solids.

The alkenes and alkynes

Alkenes are hydrocarbons with carbon-to-carbon double bonds $(R_2C=CR_2)$ and alkynes are hydrocarbons with carbon-to-carbon triple bonds $(R_{--}C \equiv C_{--}R)$. Collectively, they are called unsaturated hydrocarbons because they have fewer hydrogen atoms than does an alkane with the same number of carbon atoms.



The alkenes also form a homologous series. They have the general formula C_nH_{2n} . Alkenes end with the suffix -ene, but they can have two double bonds and then the suffix is -diene.

Ethene and propene are most often called by their common names, ethylene and propylene, respectively.



Figure 13: Ethene and propene are the simplest alkenes

Name	Molecular formula	Condensed structural formula
Ethene	$ m C_2H_4$	$\rm CH_2{=}\rm CH_2$
Propene	$\mathrm{C}_{3}\mathrm{H}_{6}$	$CH_2 = CHCH_3$
1-butene	C_4H_8	$CH_2 = CHCH_2CH_3$
1-pentene	$\mathrm{C}_{5}\mathrm{H}_{10}$	$\mathrm{CH}_2 = \mathrm{CH}(\mathrm{CH}_2)_2\mathrm{CH}_3$
1-hexene	$\mathrm{C}_{6}\mathrm{H}_{12}$	$\mathrm{CH}_2 = \mathrm{CH}(\mathrm{CH}_2)_3\mathrm{CH}_3$
1-heptene	$ m C_7H_{14}$	$\mathrm{CH}_2 = \mathrm{CH}(\mathrm{CH}_2)_4\mathrm{CH}_3$
1-octene	C_8H_{16}	$\mathrm{CH}_2{=}\mathrm{CH}(\mathrm{CH}_2)_5\mathrm{CH}_3$

Table 2: The first eight members of the alkene series

The alkenes are more reactive than the alkanes because they are unsaturated. As with the alkanes, compounds that have four or less carbon atoms are gases at room temperature. Those with five or more carbon atoms are liquids and very long chains will be solids.

Ethene is a chemical compound used in plants to stimulate the ripening of fruits and the opening of flowers. Propene is an important compound in the petrochemicals industry. It is used to make polypropylene and is also used as a fuel gas for other industrial processes.

Ethylene is a major commercial chemical. More than half of ethylene goes into the manufacture of polyethylene, one of the most familiar plastics. Propylene is also an important industrial chemical. It is converted to plastics, isopropyl alcohol, and a variety of other products.

The alkynes

In alkynes there must be at least one triple bond between two of the carbon atoms. They are unsaturated compounds and are therefore more reactive than alkanes. Their general formula is C_nH_{2n-2} . For example, 1-butyne has the molecular formula C_4H_6 . The simplest alkyne is ethyne also known as acetylene. Many of the alkynes are used to synthesise other chemical products.

$\begin{array}{c} HO-C \equiv C-OH \\ & H \\ H-C \equiv C-C \\ H \end{array}$

Figure 14: Ethyne and propyne are the two simplest alkynes

Acetylene is the industrial name for the organic compound ethyne. The raw materials that are needed to make acetylene are calcium carbonate and coal. An important use of acetylene is in oxyacetylene gas welding. The fuel gas burns with oxygen in a torch. Because the combustion of alkenes and alkynes is exothermic an incredibly high heat is produced, which is hot enough to melt metal.

Propyne



3. Which hydrocarbon series is the least reactive? Explain why.

The <u>full solutions</u> are at the end of the unit.

The alcohols

Alcohols are an important class of compounds containing the hydroxyl functional group. There are three classes of alcohols: primary, secondary, and tertiary.

An alcohol is any organic compound where there is a hydroxyl functional group (–OH) bound to a carbon atom. The general formula for a simple alcohol is $C_nH_{2n+1}OH$. Alcohols have the suffix -ol.

The simplest and most used alcohols are methanol and ethanol.



Figure 15: Methanol and ethanol are the simplest alcohols

There are three types of alcohol molecules. The type of alcohol is determined by the position of the hydroxyl functional group.

The alcohols share several properties because of the hydroxyl group. The hydroxyl group affects the solubility of the alcohols (in water). Those with shorter carbon chains are generally more soluble, and those with longer chains are less soluble. The strong hydrogen bond (intermolecular force) the hydroxyl groups in adjacent alcohol molecules gives alcohols a higher melting point and boiling point than the hydrocarbons (which have weak London forces).

A primary alcohol is one in which the hydroxyl group (–OH) is attached to a carbon atom with at least two hydrogen atoms. This will only occur when the hydroxyl group is at the end of the molecule chain.

A secondary alcohol is one in which the hydroxyl group (-OH) is attached to a carbon with only one hydrogen atom attached. This can happen somewhere in the middle of a carbon chain.

A tertiary alcohol is one in which the hydroxyl group is attached to a carbon with no hydrogen atoms attached. This will normally mean that the hydroxyl group is joined to the same carbon atom as a branch.

Note

When naming organic compounds, the position of the number in the name shows the position of the functional group. For example, propan-1-ol shows that the hydroxyl group is at the end of the carbon chain, whilst propan-2-ol shows that the hydroxyl group is attached to the second carbon in the chain.



Alcohols can have more than one hydroxyl group; alcohols with two hydroxyl groups are called diols and alcohols with three hydroxyl groups are called triols.

The alcohols have several different uses:

- Methylated spirits is ethanol with methanol added. Methanol is toxic. If ingested it forms formic acid or formate salts, which damage the central nervous system and can cause blindness, coma, or death.
- All alcohols are toxic, but in low concentrations ethanol can be used in alcoholic drinks. Ethanol is the only alcohol used in alcoholic drinks and it is used as an industrial solvent. It is also used as a solvent in medical drugs, perfumes and as an antiseptic.
- Methanol and ethanol can both be used as a fuel, and they burn more cleanly than petrol or diesel.

Alkyl halides

Alkyl halides are hydrocarbons with one hydrogen atom replaced by a halogen atom (F, Cl, Br, I). The

alkyl is due to the fact that a hydrocarbon branched group has the suffix -yl and is one of the three hydrocarbons: alkanes, alkenes or alkynes. These alkyl groups contain one or more halogen atoms, which leads to the name alkyl halides. Our focus will be on the alkane alkyl halides also known as the haloalkanes (or halogenoalkanes).

The general formula for alkyl halides with only one halogen atom is $C_nH_{2n}+1X$ where X is any halogen atom.

~

Representations of a halomethane where X can be F, Cl, Br or I: (a) structural, (b) molecular formula.

(a)
$$H \xrightarrow{H} X H$$

 $| | | |$
 $C \xrightarrow{-} C \xrightarrow{-} C \xrightarrow{-} H$ (b) $CH_3CH(X)CH_3$ (c) C_3H_7X
 $| | | |$
 $H H H$

Representations of 2-halopropane where X can be F, Cl, Br or I: (a) structural, (b) condensed structural, (c) molecular formula.



An organic compound is always named in accordance with the longest chain of carbon atoms that contains the functional group. If the substituent is an alkyl group, it is known as a branched chain.

Some uses of haloalkanes include:

- in fire extinguishers
- as aerosol propellants
- in refrigeration
- generating foamed plastics
- solvents in dry cleaning processes (not actually dry, but no water is required).

CFC stands for chlorofluorocarbons. Due to their low toxicity and low reactivity, CFCs were widely used in refrigeration and as propellants in aerosols. However, the low reactivity means that CFCs can get into the upper atmosphere where they are degraded by UV light and damage the ozone layer.

Chloroform (CHCl₃) was used as an anaesthetic for years. However, aside from causing dizziness, fatigue and headaches, it was discovered to be toxic, often fatally so. Even non-fatal doses can cause damage to the kidneys and liver. Chloroform can sometimes be found in cough syrups, although not often anymore.

Summary

In this unit you have learnt the following:

- Organic chemistry is the branch of chemistry that deals with organic molecules. An organic molecule is one that contains carbon.
- All living organisms contain carbon. Plants use sunlight to convert carbon dioxide in the air into
 organic compounds through the process of photosynthesis. Animals and other organisms then feed
 on plants to obtain their own organic compounds. Fossil fuels are another important source of carbon.
- $\cdot \,$ It is the unique properties of the carbon atom that give organic compounds certain properties.
- The carbon atom has four valence electrons, so it can bond with four other atoms, often joining to itself resulting in long chain structures. It also forms mostly covalent bonds with the atoms that it bonds to.
- An organic compound can be represented in different ways, such as by a molecular formula, a structural formula, a semi-structural formula or a condensed structural formula.
- A functional group is a particular group of atoms within a molecule, which give it certain reaction characteristics. Organic compounds can be grouped according to their functional group.
- The hydrocarbons are organic compounds that contain only carbon and hydrogen. They can be further divided into the alkanes, alkenes and alkynes, based on the type of bonds between the carbon atoms.
- The alkanes have only single bonds between their carbon atoms and are unreactive. Simple alkanes exist as a homologous series, in which adjacent members differ by a CH_2 unit.
- The alkenes have at least one double bond between two of their carbon atoms and the alkynes have at least one triple bond between two of their carbon atoms. They are both more reactive than the alkanes.
- An alcohol is an organic compound that contains a hydroxyl group (-OH).
- The alcohols have several different uses including their use as a solvent, for medicinal purposes and in alcoholic drinks.
- The alcohols share a number of properties because of the hydroxyl group. The hydroxyl group affects the solubility of the alcohols (in water). Those with shorter carbon chains are generally more soluble, and those with longer chains are less soluble.
- A molecule is said to be saturated if it contains the maximum possible number of hydrogen (or other) atoms for that molecule. The alkanes are all saturated compounds.
- A molecule is unsaturated if it does not contain the maximum number of hydrogen atoms for that molecule. The alkenes and alkynes are examples of unsaturated molecules. If a double or triple bond is broken, more hydrogen (or other) atoms can be added to the molecule.

Unit 1: Assessment

Suggested time to complete: 20 minutes

1.

- a. Name the homologous series of: CH₃CH (CH₃) CH₂CH (CH3) CH₂CH₃
- b. Give the general formula of:



- 2. Give two possible structural formulae for the compound with a molecular formula of C_4H_{10} .
- 3.
- a. What is the difference between the alkanes, alkenes and alkynes?
- b. Give the general formula for the alkynes.
- c. Of the alkanes, alkenes and alkynes which is/are:
 - i. saturated
 - ii. unsaturated?
- d. Which series is the least reactive? Explain why.
- 4. Give the general formula for the haloalkanes with only one halogen atom.
- 5. Draw the structural formulae for:
 - a. $CH_2(Br) CH_2 CH_3$
 - b. $CH_2(F) CH_3$
 - c. $CH_3CH_2CH_2CH_2OH$
 - d. $CH_3CH_2CH(OH)CH_3$

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.



b. Molecular formula: C_5H_{12} Although the condensed formula was written with the methyl group attached to the third carbon, if we look at the formula from right to left you will see that it is attached to the second carbon and so either of the representations given below are acceptable.



- 2.
- a. Condensed structural formula: $\rm CH_3 CHCHCH_3$ Molecular formula: $\rm C_4H_8$
- b. Condensed Structural formula: ${\rm CH}_2{\rm CHCH}\,({\rm CH}_3)\,{\rm CH}_3$ Molecular formula: ${\rm C}_5{\rm H}_{10}$

Back to Exercise 1.1

Exercise 1.2

- 1. An alkene has a double bond; an alkane has single bonds only.
- 2. Unsaturated hydrocarbons have double or triple bonds and are quite reactive; saturated hydrocarbons have only single bonds and are unreactive.
- 3. The alkanes are the least reactive. The saturated bonds are less reactive than the unsaturated double or triple bonds found in alkenes and alkynes.

Back to Exercise 1.2

Unit 1: Assessment

1.

- a. All the carbon atoms are bonded to the maximum number of other atoms (four) and there are only hydrogen and carbon atoms in the molecule. Therefore this is an alkane.
- b. C_nH_{2n}
- 2.



3.

- a. Alkanes have only single bonds between carbon atoms, alkenes have at least one double bond between carbon atoms and alkynes have at least one triple bond between carbon atoms.
- b. $C_n H_{2n\mathchar`-2}$
- c.
- i. alkanes
- ii. the alkenes and alkynes
- d. The alkanes are the least reactive. The saturated bonds are less reactive than the unsaturated double or triple bonds found in alkenes and alkynes.
- 4. $C_nH_{2n}+1X$ where X is any halogen atom.



Back to Unit 1: Assessment

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Unit 2: Naming organic molecules

EMMA HARRAGE



What you should know

Before you start this unit, make sure you can:

• Understand hydrocarbons and functional groups. Refer to <u>level 4 subject outcome 5.3 unit 1</u> if you need help with this.

Introduction

Parts of the text in this unit were sourced from <u>Siyavula Physical Science Gr 12 Learner's Book, Chapter 4</u>, released under a CC-BY licence.

In this unit you will learn about isomers of hydrocarbons and how to name organic compounds according to IUPAC naming conventions.

Isomers

It is possible for two organic compounds to have the same molecular formula but a different structural formula. Look at the two organic compounds that are shown in figure 1 for example.



Figure 1: Isomers of a 4-carbon organic compound with the molecular formula C_4H_{10} ; (a) is butane and (b) is 2-methylpropane

Both butane and 2-methylpropane (isobutane) are used in camping stoves and lighters.

If you were to count the number of carbon and hydrogen atoms in each C_4H_{10} compound, you would find that they are the same. They both have the same molecular formula, but their structure is different and so are their properties. Such compounds are called isomers.

Isomers are molecules with the same molecular formula and often (though not always) with the same kinds of chemical bonds between atoms, but with the atoms arranged differently.

The isomers shown in figure 1 differ only in the location of the carbon atoms. The functional groups are the same, but butane has all four carbons in one chain, while 2-methylpropane has three carbons in the longest chain and a methyl group attached to the second carbon in the chain.

It is also possible to have *positional* isomers (see figure 2). In this case the –OH functional group can be on different carbon atoms, for example carbon 1 for pentan-1-ol, on carbon 2 for pentan-2-ol or on carbon 3 for pentan-3-ol.



$CH_3CH_2CH(OH)CH_2CH_3$

It is important to note that molecules need not have the same functional groups to be isomers. For example, ethanol and methoxymethane (in figure 3) have the same molecular formula (C_2H_6O) but different functional groups and properties. These types of isomers are functional isomers.

CH3-CH2-OH CH₃−O−CH₃ Methoxymethane Ethanol

Figure 3: Functional isomers of $\mathrm{C_2H_6O}$

Figure 2: The structural and condensed structural representations of the isomers (a) pentan-1-ol, (b) pentan-2-ol and (c) pentan-3-ol

Exercise 2.1



Match the organic compound in Column A with its isomer in Column B:

Naming organic molecules

To give compounds a name, certain rules must be followed. When naming organic compounds, the International Union of Pure and Applied Chemistry IUPAC (IUPAC) nomenclature (naming scheme) is used. This is to give consistency to the names. It also enables every compound to have a unique name, which is not possible with the common names used (for example in industry). We will first look at some of the steps that need to be followed when naming a compound, and then try to apply these rules to some specific examples.

A good general rule to follow is to start at the end (the suffix) and work backwards (from right to left) in the name.

Molecules can contain both double or triple bonds and other functional groups (e.g. an alkene and an alcohol functional group in one molecule – propenol). However, all molecules explored in this book will contain only single carbon-carbon bonds when combined with other functional groups.

1. Recognise the functional group in the compound. This will determine the suffix of the name.

Table 1: The suffix associated with various functional groups

Functional group	Suffix
Alkane	-ane
Alkene	-ene
Alkyne	-yne
Alcohol	-ol

2. Find the longest continuous carbon chain that contains the functional group (it won't always be a straight chain) and count the number of carbon atoms in this chain. This number will determine the prefix (the beginning) of the compound's name.

Table 2: The prefix of a compound's name is determined by the number of carbon atoms in the longest chain that contains the functional group

Number of carbon atoms	Prefix
1	Meth-
2	Eth-
3	Prop-
4	But-
5	Pent-
6	Hex-
7	Hept-
8	Oct-
9	Non-
10	Dec-

- 3. Number the carbons in the longest carbon chain containing the functional group. (Important: if the molecule is not an alkane i.e. has a functional group, you need to start numbering so that the functional group is on the carbon with the lowest possible number). Start with the carbon at the end closest to the functional group.
- 4. Look for any branched groups:

Name them by counting the number of carbon atoms in the branched group and referring to Table 2, these groups will all end in -yl.

Note the position of the group on the main carbon chain. If there is more than one of the same types of branched group, then both numbers must be listed (e.g. 2,4 -) and one of the prefixes listed in Table 3 below must be used. (Important: if the molecule is an alkane the branched group must be on the carbon with the lowest possible number.)

The branched groups must be listed before the name of the main chain in alphabetical order (ignoring di/tri/tetra).

If there are no branched groups, this step can be ignored.

Table 3: Prefixes for multiple substituents with the same name; these apply to multiple functional groups as well

Number	Prefix
2	Di-
3	Tri-
4	Tetra-

For the alkyl halides the halogen atom is treated in much the same way as branched groups:

To name them, take the name of the halogen atom (e.g. iodine) and replace the -ine with -o (e.g. iodo).

Table 4: Naming halogen atoms in organic molecules

Halogen	Name
Fluorine	Fluoro
Chlorine	Chloro
Bromine	Bromo
lodine	lodo

Give the halogen atom a number to show its position on the carbon chain. If there is more than one halogen atom the numbers should be listed and a prefix should be used (e.g. 3,4-diiodo- or 1,2,2-trichloro-). See Table 4 for a list of the prefixes.

The halogen atoms must be listed before the name of the main chain in alphabetical order (ignore di/ tri/tetra).

If there are no halogen atoms this step can be ignored.

- 5. Combine the elements of the name into a single word in the following order:
 - branched groups/halogen atoms in alphabetical order (ignoring prefixes)
 - prefix of main chain
 - finally, name the ending according to the functional group and its position on the longest carbon chain.



Solution

1. Identify the functional group

The compound is a hydrocarbon with single bonds between the carbon atoms. It is an alkane and will have a suffix of -ane.

2. Find the longest carbon chain

There are four carbon atoms in the longest chain. The prefix of the compound will be but-.

3. Number the carbon atoms in the longest chain

The numbering has been done for you here.

4. Look for any branched group, name them and give their position on the carbon chain

There are no branched groups in this compound.

5. Combine the elements of the name into a single word

The name of the compound is butane.



Example 2.2

Name this alkanes:



Solution

1. Identify the functional group

The compound is a hydrocarbon with single bonds between the carbon atoms. It is an alkane and will have the suffix -ane.

2. Find the longest carbon chain

There are three carbon atoms in the longest chain. The prefix for this compound is prop-.
3. Number the carbons in the carbon chain

If we start at the carbon on the left, we can number the atoms as shown in red (below left). If we start at the carbon on the right, we can number the atoms as shown in blue (below right).



4. Look for any branched groups, name them and give their position on the carbon chain There is a branched group attached to the second carbon atom. In this case the methyl group is on carbon 2 regardless of which side you number the longest chain from.

This group has the formula CH_3 , which is methane without a hydrogen atom. However, because it is not part of the main chain, it is given the suffix -yl (i.e. methyl). The position of the methyl group comes just before its name (see the next step).

5. Combine the elements of the compound's name into a single word in the order of branched group; prefix; name ending according to the functional group

The compound's name is 2-methylpropane.



Example 2.3

Give the IUPAC name for the following compound:

 $\mathrm{CH}_3\mathrm{CH}\,(\mathrm{CH}_3)\,\mathrm{CH}\,(\mathrm{CH}_3)\,\mathrm{CH}_3$

Solution

Remember that the side groups are shown in brackets after the carbon atom to which they are attached.

1. Draw the structural formula from its condensed structural formula:



2. Identify the functional group

The compound is a hydrocarbon with single bonds between the carbon atoms. It is an alkane and will have the suffix -ane.

3. Find the longest carbon chain

There are four carbon atoms in the longest chain. The prefix for this compound is but-.

4. Number the carbons in the carbon chain

If we start at the carbon on the left, we can number the atoms as shown in red (below left). If we start at the carbon on the right, we can number the atoms as shown in blue (below right).



5. Look for any branched groups, name them and give their position on the carbon chain

There are two methyl groups attached to the main chain. The first one is attached to the second carbon atom and the second methyl group is attached to the third carbon atom. Notice that in this example it does not matter how you have chosen to number the carbons in the main chain; the methyl groups are still attached to the second and third carbon atoms and so the naming of the compound is not affected.

This group will be 2,3-dimethyl-

6. Combine the elements of the compound's name into a single word in the order of branched groups; prefix; name ending according to the functional group

The compound's name is 2,3-dimethylbutane.





There is one methyl group attached to the main chain. If we number as shown in red (on the left) the methyl is attached to the fourth carbon atom. If we number as shown in blue (on the right) the methyl is attached to the third carbon atom.

- After functional groups, the branched groups should have the lowest numbers possible. Therefore the blue numbering (on the right) is correct. The methyl is attached to the third carbon atom (3-methyl).
- 5. Combine the elements of the compound's name into a single word in the order of branched groups; prefix; name ending according to the functional group

The compound's name is 3-methylhexane.



Example 2.5

Draw the semi-structural, structural and condensed structural formula for the organic compound 2,2,4-trimethylhexane.

Solution

1. Identify the functional group

The name ends in -ane therefore the compound is an alkane.

2. Determine the number of carbon atoms in the longest chain

The longest chain has the prefix hex-. There are therefore 6 carbon atoms in the longest chain.

Structural formula:

$$c - c - c - c - c - c$$

3. Look for any branched groups and place them on the structure. The compound is 2,2,4-trimethylhexane. Therefore there are three branched groups. Two on carbon 2 and one on carbon 4.

Semi-structural formula:



4. Combine this information and add the hydrogen atoms

Carbon atoms can have four single bonds. Therefore wherever a carbon atom has less than four bonds draw in hydrogen atoms until there are four bonds.



5. Condense the structural formula

First condense the main chain: $\rm CH_3 \rm CCH_2 \rm CHCH_2 \rm CH_3$

Then add the side chains (in brackets) on the relevant carbon atoms:

Condensed structural formula: CH₃C(CH₃)₂CH₂CH (CH₃) CH₂CH₃



Exercise 2.2

- 1. Give the structural formula for each of the following alkanes:
 - a. octane
 - b. 3-ethylpentane
- 2. Give the IUPAC name for each of the following alkanes:
 - a. $CH_3CH_2CH(CH_3)CH_2CH_3$
 - b. $CH_3CH(CH_3)CH_2CH(CH_3)CH_3$
 - c.



Naming alkenes

Let's look at some examples of how to name alkenes.





Solution

1. Identify the functional group

The compound has a double carbon-carbon bond and is an alkene. It will have the suffix -ene.

2. Find the longest carbon chain containing the functional group

The functional group is a double bond, so the longest chain must contain the double bond. There are four carbon atoms in the longest chain and so the prefix for this compound will be but-.

3. Number the carbon atoms

Remember that the carbon atoms must be numbered so that the functional group is at the lowest numbered carbon atom possible. In this case, it doesn't matter whether we number the carbons from the left to right, or from the right to left. The double bond will still fall between the second and third carbon atoms.

4. Look for any branched groups, name them and give their position on the carbon chain

There are no branched groups in this molecule.

5. Combine the elements of the name into a single word in the following order: branched groups; prefix; name ending according to the functional group and its position along the longest carbon chain

The name of this compound is but-2-ene or 2-butene.



Example 2.7

Draw the structural and molecular formula for the organic compound 3-methylbut-1-ene.

Solution

1. Identify the functional group

The suffix -ene means that this compound is an alkene and there must be a double bond in the molecule. The number 1 immediately before the suffix means that the double bond must be at the first carbon in the chain (but-1-ene).

2. Determine the number of carbon atoms in the longest chain containing the functional group





Solution

- Identify the functional group The compound is an alkene and will have the suffix -ene. There is a double bond between the first and second carbons and between the third and fourth carbons. The organic compound therefore contains '1,3-diene'.
- 2. Find the longest carbon chain containing the functional group, and number the carbon atoms Remember that the main carbon chain must contain both the double bonds. There are four carbon atoms in the longest chain containing the double bonds and so the prefix for this compound will be but-. The carbon atoms are already numbered 1 to 4 in the diagram.
- Look for any branched groups, name them and give their position on the carbon chain There is an ethyl group on the second carbon. Note that if we had numbered from the right to left the suffix would still have been 1,3-diene, however the ethyl group would have been on the third carbon. So we have to number from left to right.
- 4. Combine the elements of the name into a single word in the following order: branched groups; prefix; name ending according to the functional group and its position along the longest carbon chain

The name of this compound is 2-ethylbut-1,3-diene.



Exercise 2.3

- 1. Give the IUPAC name for each of the following alkenes:
 - a. $CH_2CHCH_2CH_2CH_3$

b.



Naming alkynes

The suffix for an alkyne is -yne. Let's look at some examples of how to name alkynes.



In this example, you will need to number the carbons from right to left so that the triple bond is between carbon atoms with the lowest numbers (the suffix for the compound will therefore be -2-yne).



- Look for any branched groups, name them and assign the number of the carbon atom to which the group is attached.
 There is a methyl (CH₃) group attached to the fifth carbon (remember we have numbered the carbon atoms from right to left).
- 5. Combine the elements of the name into a single word in the following order: branched groups; prefix; name ending according to the functional group and its position along the longest carbon chain

If we follow this order, the name of the compound is 5-methylhex-2-yne.





- Numbering from left to right (shown in red) the first triple bond is on carbon 1 and the second is on carbon 5. The suffix will therefore be -1,5-diyne. (Numbering from right to left (shown in blue) will give the suffix -2,6-diyne, and is incorrect).
- 5. Look for any branched groups There are no branched groups for this molecule.
- 6. Combine the elements of the name into a single word in the following order: branched groups; prefix; name ending according to the functional group and its position along the longest carbon chain

The name of the compound is hept-1,5-diyne.





The <u>full solutions</u> are at the end of the unit.

Naming alkyl halides

All the same rules apply when naming the alkyl halides as for naming the hydrocarbons. The halogen atom is treated in the same way as a branched group.



Example 2.13



Solution

- Identify the functional group There are three halogen atoms and no other functional groups. This compound is therefore a haloalkane and will have the suffix -ane.
- 2. Find the longest carbon chain containing the functional group There are four carbons in the longest chain containing all the halogen atoms. The prefix for this compound will be but-.
- 3. Number the carbon atoms in the longest chain You need to number the carbon atoms so that the halogen atoms are on the carbon atoms with the lowest numbers. You must number from left to right here so that one halogen atom is on carbon 1 and two halogen atoms are on carbon 3.
- Name the halogen atoms and assign the number for the carbon atom attached to it There are two halogen atoms that are bromine atoms and one that is fluorine. One bromine is attached to carbon 1 and one is attached to carbon 3. The fluorine atom is attached to carbon 3. So you have 1,3-dibromo- and 3-fluoro.
- 5. Look for branched groups There are no branched groups in this compound.
- 6. Combine the elements of the name into a single word in the following order: halogen atoms in alphabetical order; prefix; name ending according to functional group The name of the compound is 1,3-dibromo-3-fluorobutane. Note that we place the halogens in alphabetical order: bromo (ignore the di/tri/tetra) is before fluoro.

Example 2.14

Draw the structural and condensed structural formula for the organic compound 2-iodo-3-methylpentane.

Solution

- Identify the functional group This compound has the suffix -ane, but also contains a halogen atom. It is therefore a haloalkane. Note that the methyl and iodo are written in alphabetical order.
- 2. Find the longest carbon chain containing the functional group



Naming alcohols

The rules used to name the alcohols are like those already discussed for the hydrocarbons. The suffix of an alcohol is -ol.



Give the IUPAC name for the following organic compound:



Solution

- Identify the functional group The compound has an -OH (hydroxyl) functional group and is therefore an alcohol.
- 2. Find the longest carbon chain containing the functional group There are three carbon atoms in the longest chain that contains the functional group. The prefix for this compound will be prop-. As there are only single bonds between the carbon atoms, the prefix includes an- to become propan-.
- Number the carbons in the carbon chain In this case, it doesn't matter whether you start numbering from the left or right. The hydroxyl group will still be attached to second carbon atom (-2-ol).
- Look for branched groups
 There are no branched groups in this compound.
- 5. Combine the elements of the compound's name into a single word in the order of branched groups; prefix; name ending according to the functional group The compound's name is propan-2-ol or 2-propanol.



Example 2.16

Give the IUPAC name for the following compound:



1. Identify the functional group

The compound has an -OH (hydroxyl) functional group and is therefore an alcohol. There are two hydroxyl groups in the compound, so the suffix will be -diol.

- 2. Find the longest carbon chain that contains the functional group There are four carbon atoms in the longest chain that contains the functional group (but-) and only single bonds (an-). The prefix for this compound will be butan-.
- 3. Number the carbons in the carbon chain:



There are two hydroxyl groups attached to the main chain. If we number as shown in red (on the left) they are attached to the first and second carbon atoms. If we number as shown in blue (on the right) they are attached to the third and fourth carbon atoms.

- 4. The functional groups should have the lowest numbers possible. Therefore the red numbering is correct. The hydroxyl groups are attached to the first and second carbon atoms (1,2-diol).
- 5. Look for branched groups There are no branched groups in this compound.
- 6. Combine the elements of the compound's name into a single word in the order of branched groups; prefix; name ending according to the functional group The compound's name is butan-1,2-diol.



Example 2.17

Draw the structural and condensed structural representations for the organic compound 4-ethyloctan-2,5-diol.

Solution

- Identify the functional group The compound has the suffix -ol. It is therefore an alcohol.
- Find the longest carbon chain that contains the functional group The prefix is oct- therefore there are 8 carbons in the longest chain containing the functional group.



3. Place the functional group as well as any branched groups There is one -OH attached to carbon 2 and one attached to carbon 5. There is also an ethyl ($-CH_2CH_3$) branched group attached to carbon 4.



Summary

In this unit you have learnt the following:

- If two compounds are isomers, it means that they have the same molecular formula but different structural formulae.
- · Organic compounds are named according to their functional group and its position in the molecule,

the number of carbon atoms in the molecule and the position of any double and triple bonds. The IUPAC rules for nomenclature are used in the naming of organic molecules.

Unit 2: Assessment

Suggested time to complete: 30 minutes

- 1. Give the structural formula for each of the following compounds:
 - a. 3-methylhexane
 - b. 4-ethyl-4-methylhept-2-yne
 - c. 2,3-difluoropentane
 - d. 2-methylpropan-1-ol
- 2. Give the IUPAC name for the following haloalkanes:



c. $CH_2(F) C(I)_2 CH_2 CH_3$

3. Give the IUPAC name for each of the following:

a.
$$CH_3CH_2CH(OH)CH_3$$

b.



The <u>full solutions</u> are at the end of the unit.

Unit 2: Solutions

Exercise 2.1

- 1. Al (CH₃CH₂CH₂ (OH)) and B3 (CH₃CH (OH) CH₃) are isomers.
- 2. B1 $(CH_3CH(CH_3)CH_3)$ and A2 are isomers.
- 3. B2 and A3 are isomers

Back to Exercise 2.1

Exercise 2.2

1.

a. The suffix -ane tells us that this is an alkane. The prefix oct- tells us that there are eight carbon atoms in the longest chain.



b. The suffix -ane tells us that this is an alkane. The prefix pent- tells us that there are five carbon atoms in the longest chain. 3-ethyl tells us that there is an ethyl branched group ($\rm CH_3CH_2$ -) group attached to the third carbon atom.



2.

a. This is hard to do unless you draw the structural formula of the molecule. It is recommended that you do this in exams.

There are five carbons in the longest chain, so the prefix is pent-. There are only single carboncarbon bonds and no other functional group, so the compound is an alkane, and the suffix is -ane. There is one methyl group at position 3 (you can number from either end of the chain for this example). So the compound is 3-methylpentane.



b. There are five carbons in the longest chain, so the prefix is pent-. There are only single carboncarbon bonds, so the compound is an alkane, and the suffix is -ane. There is one methyl group at position 2 and one at position 4 (once again you can number from either end). So the compound is 2,4-dimethylpentane.



- c. There are four carbon atoms in the longest chain, so the prefix is but-. There are two methyl branches at positions 2 and 3. The functional group is an alkane, so the suffix is -ane. Combining all this information we get: 2,3-dimethylbutane. Note that in this example *it does not matter* which end you start numbering from.
- d. There are four carbon atoms in the longest chain, so the prefix is but-. The functional group is an alkane, so the suffix is -ane. There is one branched group which is a methyl group, and this is at position 2. The molecule is 2-methylbutane.

Note that in this example *it does matter* which way your number the chain as the branched group needs to have the lowest number possible and so the compound is not 3-methylbutane.

Back to Exercise 2.2

Exercise 2.3

٦.





The molecule contains a double carbon-carbon bond. It is an alkene and so the suffix is -ene. There are five carbons in the longest chain, so the prefix is pent-. There are no branched groups. The double bond occurs between carbons 1 and 2. So the molecule is 1-pentene or pent-1-ene.

Note that the way you number the carbon atoms matters here, the molecule is not pent-4-ene.

b. The compound contains two double carbon-carbon bonds. It is an alkene and so the suffix is
 -diene. There are four carbons in the longest chain containing the double bonds, so the prefix is
 but- there are no branched chains. The first double bond occurs between carbons one and two.
 The second double bond occurs between carbons three and four. The compound is but-1,3-diene.

a. The prefix hex- tells us there are six carbon atoms in the chain. The suffix -1-ene tells us there is a double bond between the first and second carbon atoms.



b. The prefix oct- tells us there are eight carbon atoms in the longest chain containing the functional group. The suffix -3-ene tells us there is a double bond between the third and fourth carbon atoms.
 4-ethyl tells us that there is an ethyl (CH₃CH₂-) branched group attached to the fourth carbon atom.



Back to Exercise 2.3

Exercise 2.4

1.

a. The prefix pent- tells us there are five carbon atoms in the longest chain. The suffix -1-yne tells us there is a triple bond between the first and second carbon atoms.



b. The prefix hept- tells us there are seven carbon atoms in the longest chain. The suffix -3-yne tells us there is a triple bond between the third and fourth carbon atoms. 5-methyl tells us there is a methyl branched chain on the fifth carbon atom.



2.

a. There is a triple carbon-carbon bond. This compound is an alkyne and will have the suffix -yne. There are four carbon atoms in the longest chain, therefore the prefix will be but-. The triple bond is between the second and third carbon atoms regardless of how you number the chain (-2-yne). There are no branched groups. This molecule is but-2-yne or 2-butyne.



b. There is a triple carbon-carbon bond. This compound is an alkyne and will have the suffix -yne. There are six carbon atoms in the longest chain, therefore the prefix will be hex-. The triple bond is between the third and fourth carbon atoms regardless of how you number the chain (-3-yne). There are two branched methyl groups. Depending on the order of numbering they are either on the fourth carbon atom (left to right) or the second carbon atom (right to left). The lower numbering is correct (right to left) and as there are two methyl groups this is 2,2-dimethyl. The molecule is 2,2-dimethylhex-3-yne.

Back to Exercise 2.4

Unit 2: Assessment

1.

a. The prefix hex- tells us there are six carbon atoms in the longest chain. The suffix -ane tells us this is an alkane and that there are only single carbon-carbon bonds and no other functional groups.
3-methyl tells us there is a branched methyl group on the third carbon atom.



b. The prefix hept- tells us there are seven carbon atoms in the longest chain. The suffix -2-yne tells us this is an alkyne and there is a triple bond between the second and third carbon atoms. The 4-ethyl tells us there is an ethyl (CH_3CH_2 -) branched group on the fourth carbon atom. The 4-methyl tells us there is a methyl (CH_3 -) branched group on the fourth carbon atom.



c. The pent- tells us there are five carbon atoms in the longest chain. The -ane tells us there are only single carbon-carbon bonds. The 2,3-difluoro means there are two fluorine atoms, one attached to carbon 2 and the other attached to carbon 3.



d. The prefix propan- tells us there are three carbon atoms in the longest chain and only single carbon-carbon bonds. The suffix -1-ol tells us there is a hydroxyl group attached to the first carbon atom. 2-methyl tells us there is a methyl branched group attached to the second carbon atom.



2.

- a. There is a halogen atom and only single carbon-carbon bonds, therefore this is a haloalkane and the suffix is -ane. There are three carbon atoms in the longest chain, therefore the prefix is prop-. There is an iodine atom attached to the second carbon atom (2-iodo). There are no branched groups. Therefore this molecule is 2-iodopropane.
- b. There are halogen atoms and only single carbon-carbon bonds, therefore this is a haloalkane and the suffix is -ane. There are eight carbon atoms in the longest chain, therefore the prefix is oct-. There are two chlorine atoms, one attached to the first carbon atom, and one attached to the fourth carbon atom (1,4-dichloro). There are no branched groups. This molecule is 1,4-dichlorooctane.

c.



There are halogen atoms and only single carbon-carbon bonds, therefore this is a haloalkane and the suffix is -ane. There are four carbon atoms in the longest chain containing all the halogen atoms, therefore the prefix is but-. There is a fluorine atom attached to the first carbon atom (1-fluoro). There are two iodine atoms attached to the second carbon atom (2,2-diiodo). These must be put in alphabetical order, ignoring any prefixes on the halogen atoms. F comes before i (1-fluoro-2,2-diiodo). This molecule is 1-fluoro-2,2-diiodobutane.

3.

a.



There is a hydroxyl group, therefore the compound is an alcohol, and the suffix is -ol. There are four carbon atoms in the longest chain, so the prefix is but-. There are only single carbon-carbon bonds, therefore the prefix becomes butan-. There are no branched groups. The hydroxyl group is attached to the second carbon atom. The molecule is 2-butanol or butan-2-ol.

Note that the way we number the carbon atoms matters. The hydroxyl group is given the lowest possible number and so this compound is not butan-3-ol.

b. There are two hydroxyl groups, therefore the compound is an alcohol, and the suffix is -diol. There are seven carbon atoms in the longest chain, so the prefix is hept-. There are only single carbon-carbon bonds, therefore the prefix becomes heptan-. The hydroxyl groups are both attached to the third carbon atom (-3,3-diol). The branched methyl chain is attached to the fourth carbon atom. The molecule is 4-methylheptan-3,3-diol.

Note that the way we number the carbon atoms matters. The hydroxyl groups are given the lowest possible numbers and so this compound is not 4-methylheptan-5,5-diol.

Back to Unit 2: Assessment

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Unit 3: Physical Properties of Organic molecules

EMMA HARRAGE



What you should know

Before you start this unit, make sure you:

- Understand hydrocarbons and functional groups. Refer to <u>level 4 subject outcome 5.3 unit 1</u> if you need help with this.
- Can name organic molecules. Refer to level 4 subject outcome 5.3 unit 2 if you need help with this.
- Understand polarity. Refer to level 3 subject outcome 5.2 unit 5 if you need help with this.
- Can recall the different types of bonding in covalent molecules. Refer to <u>level 3 subject outcome 5.3</u> <u>unit 1</u> if you need help with this.

Introduction

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In this unit you will learn about the chemical and physical properties of hydrocarbons. This will include density, melting and boiling points, and how chain length affects these properties.

Physical properties

The types of intermolecular forces that occur in a substance will affect its physical properties, such as its phase, melting point, and boiling point. You should remember from the kinetic theory of matter that the phase of a substance is determined by how strong the forces are between its particles. The weaker the forces, the more likely the substance is to exist as a gas. This is because the particles can move far apart since they are not held together very strongly. If the forces are very strong, the particles are held closely together in a solid structure. Remember also that the temperature of a material affects the energy of its particles. The more energy the particles have, the more likely they are to be able to overcome the forces that are holding them together. This can cause a change in phase.

The physical properties of any compound depend largely on:

- 1. its mass
- 2. its type of intermolecular and intramolecular forces of attraction

3. its melting point and boiling point.

Compounds that contain very similar atoms can have very different properties depending on how those atoms are arranged. This is especially true when they have different functional groups. Table 1 shows some properties of different homologous series.

Homologous series	Typical smell	Example	Formula	Melting point (°C)	Boiling point (°C)	Phase (at 25 °C)
alkane	odourless	ethane	$\mathrm{C}_{2}\mathrm{H}_{6}$	-183	-89	gas
alkene	sweet/musky	ethene	$\mathrm{C}_{2}\mathrm{H}_{4}$	-169.2	-103.7	gas
haloalkane	almost odourless	chloro ethane	$\mathrm{C}_{2}\mathrm{H}_{5}\mathrm{Cl}$	-139	12.3	gas
alcohol	sharp	ethanol	C_2H_6O	-114	78.4	liquid
alkyne	odourless	ethyne	$\mathrm{C}_{2}\mathrm{H}_{2}$	-80.8	-84	gas

Table 1: The properties of some homologous series

Listed in the table are the common smells and other physical properties for common homologous series. Only one representative example from each homologous series is provided. This does not mean that all compounds in that series have the same properties. For example, short-chain and long-chain alkanes are generally odourless, while those with moderate chain length (approximately 6 - 12 carbon atoms) smell like petrol.

Intermolecular forces

Intermolecular forces affect the boiling and melting points of substances. Substances with weak intermolecular forces will have low melting and boiling points as less energy (heat) is needed to overcome these forces. Those with strong intermolecular forces will have high melting and boiling points as more energy (heat) is required to overcome these forces. When the temperature of a substance is raised beyond its melting or boiling point the intermolecular forces are not weakened. Rather, the molecules have enough energy to overcome those forces.

Table 2: The relationship between intermolecular forces and melting points, boiling points, and physical states

Name	Main Intermolecular forces	Molecular mass $(g.mol^{-1})$	Melting point (°C)	Boiling point (°C)	Phase (at)
ethane	induced-dipole	30.06	-183	-89	gas
dimethyl ether	dipole-dipole	46.06	-141	-24	gas
chloroethane	dipole-dipole	64.5	-139	12.3	gas
pentane	induced-dipole	72.12	-130	36	liquid
propan-1-ol	hydrogen bonds	60.08	-126	97	liquid
ethanol	hydrogen bonds	46.06	-114	78.4	liquid
butan-1-ol	hydrogen bonds	74.1	-90	118	liquid

As the strength of the intermolecular forces increases (from top to bottom in table 2) the melting and boiling points increase. The stronger the intermolecular forces the more likely a substance is to be a liquid or a solid at room temperature.

Flammability

Flammability is a measure of how easy it would be for a substance to catch alight and burn. The flash point of a substance is the lowest temperature that is likely to form a gaseous mixture you could set alight. If a liquid has a low enough flash point it is considered flammable (able to be ignited easily) while those with higher flash points are considered non-flammable. A substance that is classified as non-flammable can still be forced to burn, but it will not ignite easily.

Table 3: The relationship between	intermolecular	forces and	the flamma	bility of
	a substance			-

Name	Main intermolecular forces	Vapour pressure (kPa at $20\ ^\circ { m C}$)	Flash point (°C)	Flammability
ethane	induced-dipole	3 750	-135	very high
propane	induced-dipole	843	-104	very high
dimethyl ether	dipole-dipole	510	41	very high
butane	induced-dipole	204	-60	very high
chloroethane	dipole-dipole	132.4	-50	very high
pentane	induced-dipole	57.9	-49	very high
propanone	dipole-dipole	24.6	-17	high
ethanol	hydrogen bonds	5.8	17	high
water	hydrogen bonds	2.3	-	very low
propan-1-ol	hydrogen bonds	2	22	high
butan-1-ol	hydrogen bonds	0.6	35	high

When a substance is in a liquid or solid state there will be some molecules in the gas state. These molecules have enough energy to overcome the intermolecular forces holding most of the substance in the liquid or solid phase. These gas molecules exert a pressure on the liquid or solid (and the container) and that pressure is the vapour pressure of that compound. The weaker the intermolecular forces within a substance the higher the vapour pressure will be. Compounds with higher vapour pressures have lower flash points and are therefore more flammable.



boiling point: 100 °C



boiling point: 78 °C



Figure 1: The relative vapour pressures of water, ethanol and propane



Use the information in the table below to answer the following questions.

	Name	Boiling point (°C)
А	Ethyne	-84.0
В	Ethyne	-89.0
С	Ethanol	78.4

- 1. Which of the compounds listed in the table are gases at room temperature?
- 2. Name the main type of intermolecular forces for A, B and C.
- 3. Briefly explain the difference in boiling point between B and C.
- 4. Draw the structural representations of A, B and C.

The <u>full solutions</u> are at the end of the unit.

The physical properties of alkanes

Alkanes are the simplest type of hydrocarbons. There are only C-H bonds throughout the whole structure of alkanes. Alkanes do not have any functional group. As the number of carbon atoms increases, the size of the molecule also increases.

As there are a greater number of atoms, there are a greater number of electrons, which means stronger intermolecular forces. So as the number of carbon atoms in alkanes increase, their melting and boiling points also increases. Alkanes with low molecular mass such as methane, ethane and propane are gas at room temperature, alkanes with greater molecular mass such as octane are liquid at room temperatures.

C-H bonds are non-polar, which means there are only weak London forces between molecules. So most alkanes have a low melting and boiling points. Most alkanes are usually volatile, they easily evaporate.

Alkanes are completely insoluble in water, because in water there is stronger hydrogen bonding.

The physical properties of alkenes

Alkenes have a carbon-to-carbon double bond in their structure (C=C). This gives them a unique chemical property and similar trends in physical properties. There are also C-H bonds throughout their structure. Both C=C and C-H bonds are non-polar, which means there are no strong dipole-dipole interactions present between molecules. There are only weak London forces. Therefore alkenes have low melting and boiling points. Most of the short chain alkenes are usually gases at room temperature.

Again, as the carbon number increases, the molecule gets bigger, which means more electrons and stronger intermolecular forces between the molecules. So as the number of carbon atoms in the chain increases, the melting and boiling points of the alkenes also increases.

Alkenes are also insoluble in water. Like alkanes, there are only weak London forces between the alkene molecules, while water molecules have a much stronger hydrogen bonding between them. Hence, they do not mix.

The physical properties of alkynes

The properties of alkynes pretty much follow the same pattern of those of alkanes and alkenes. Alkynes are unsaturated hydrocarbons that have a $C \equiv C$. All alkynes are odourless and colourless except for ethylene which has a slight distinctive odour.

The first three alkynes are gases, and the next eight are liquids. All alkynes with longer chains than these eleven are solids.

The boiling point and melting point of alkynes increases as their molecular mass increases. The boiling points of alkynes are slightly higher than those of their corresponding alkenes, because of one extra bond at the carbon site.

The physical properties of alcohols

Alcohols have an -OH group present in their structure, which is highly polar, which makes the alcohol molecules polar too. Alcohols are so polar, that they form very strong hydrogen bonds between molecules which are much stronger than London forces. Therefore alcohols have much higher melting and boiling points when compared to both alkanes and alkenes. Most alcohols are usually liquid at room temperature.



Figure 2: Hydrogen bonding in methanol

Alcohols are highly soluble in water, because both water and alcohols can form hydrogen bonds with each other.

The physical properties of haloalkanes

As the chain length increases in the homologous series of the haloalkanes group, differences can be seen because of the difference in atomic mass of the compounds.

There is a large electronegativity difference between a halogen atom and a carbon atom, resulting in highly polarised molecules. The higher molecular mass and greater polarity as compared to the parent
hydrocarbon molecule results in stronger intermolecular forces of attraction: dipole-dipole and van der Waals. Boiling points depend upon the intermolecular forces of attraction and hence the boiling points of chloro-alkanes, bromo-alkanes and iodo-alkanes are considerably higher than those of the hydrocarbons of comparable molecular mass.

As the chain length increases in the homologous series of haloalkanes, the forces of attraction become stronger due to the increase in molecular size and mass, hence the boiling point increases. But, as with all hydrocarbons, the boiling point decreases with branching.

chloromethane, bromomethane, chloroethane and some chlorofluoromethanes are gases at room temperature. Higher members are liquids or solids.

The melting point of a compound depends upon the strength of the intermolecular forces between the molecules. The melting point also follows the same trend as the boiling point.

Haloalkanes are slightly soluble in water. This is because of the relatively larger amount of energy required to break the bond between halogen and carbon and the smaller amount of energy released when the bond is formed after the dissolution of the ion and water. The more halogens that are substituted, the less volatile the haloalkane becomes. Take for example the haloalkane series shown in figure 3.



Figure 3: A series of haloalkanes with increasing numbers of chlorine atoms: (a) chloromethane, (b) dichloromethane, (c) trichloromethane and (d) tetrachloromethane

Physical properties and chain length

Remember that the alkanes are a group of organic compounds that contain carbon and hydrogen atoms bonded together. The carbon atoms link together to form chains of varying lengths. Remember that the alkanes are relatively unreactive because of their stable C-C and C-H bonds. The boiling points and melting points of these molecules are determined by their molecular structure and their surface area.

The more carbon atoms there are in an alkane, the greater the surface area available for intermolecular interactions.

H	H H	ННН
H-C-H	H - C - C - H	H - C - C - C - H
Н	нн	й й й
Methane	Ethane	Propane
$\begin{array}{cccccc} H & H & H & H \\ & & & & & & \\ H - C - C - C - C - C - H \\ & & & & & & \\ H & H & H & H \end{array}$	$\begin{array}{cccccccccccccccccccccccccccccccccccc$	$\begin{array}{cccccccccccccccccccccccccccccccccccc$
Dutane	Pentane	Hexane
$\begin{array}{cccccccccccccccccccccccccccccccccccc$	$\begin{array}{cccccccccccccccccccccccccccccccccccc$	$\begin{array}{cccccccccccccccccccccccccccccccccccc$
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Heptane	Octane	Nonane

Figure 4: Van der Waals intermolecular forces increase as chain length increases



Formula	Name	Molecular mass (g.mol ⁻¹)	Melting point ($^{\circ}\mathrm{C}$)	Boiling point (° $_{ m C}$)	Phase (at $25~^\circ\mathrm{C}$)
CH_4	methane	16.04	-182	-162	gas
$\mathrm{C}_{2}\mathrm{H}_{6}$	ethane	30.06	-183	-89	gas
C_3H_8	propane	44.08	-188	-42	gas
$ m C_4H_{10}$	butane	58.1	-137	0	gas
$\mathrm{C}_{6}\mathrm{H}_{14}$	hexane	86.14	-95	68.5	liquid
C_8H_{18}	octane	114.18	-57	125.5	liquid
$\mathrm{C}_{20}\mathrm{H}_{42}$	icosane	282.42	37	343	solid

Table 4:	The physica	l properties	of alkanes

Notice that when the molecular mass of the alkanes is low (i.e. there are few carbon atoms), the organic compounds are gases because the intermolecular forces are weak. As the number of carbon atoms and the molecular mass increases, the compounds are more likely to be liquids or solids because the intermolecular forces are stronger.

The larger a molecule is, the stronger the intermolecular forces are between the molecules. This is one of the reasons why methane is a gas at room temperature while hexane is a liquid and icosane is a solid.

Be careful when comparing molecules with different types of intermolecular forces. For example, small molecules with hydrogen bonding can have stronger intermolecular forces than large molecules with only van der Waals forces.

Density also increases with molecular size. Flash point increases with increasing chain length meaning that the longer chains are less flammable although they can still be ignited.

It is partly the stronger intermolecular forces that explain why petrol (mainly octane) is a liquid, while candle wax $(C_{23}H_{48})$ is a solid. If these intermolecular forces did not increase with increasing molecular size, we would not be able to put liquid fuel into our cars or use solid candles.



Exercise 3.2

Refer to the data table below which shows the melting point and boiling point for several organic compounds with different functional groups.

Formula	Name	Melting point (°C)	Boiling point (°C)
C_4H_{10}	Butane	-137	0
C_5H_{12}	Pentane	-130	36
C_6H_{14}	Hexane	-95	68.5
C_4H_8	But-1-ene	-185	-6.5
C_5H_{10}	Pent-1-ene	-165.2	30
C_6H_{12}	Hex-1-ene	-140	63

- 1. At room temperature (approx. $25~^\circ\mathrm{C}$.), which of the organic compounds in the table are gases?
- 2. In the alkanes, describe what happens to the melting point and boiling point as the number of carbon atoms in the compound increases. Explain your answer.

The <u>full solutions</u> are at the end of the unit.

Chemical properties

The term 'chemical properties' describes the way that an element or compound reacts chemically with other substances. A chemical property is a characteristic or behaviour of a substance that may be observed when it undergoes a chemical change or reaction.

Chemical properties are seen either during or following a reaction since the arrangement of atoms within a sample must be disrupted for the property to be investigated. This is different from a physical property, which is a characteristic that may be observed and measured without changing the chemical identity of a specimen.

Examples of chemical properties include flammability, toxicity, chemical stability, and heat of combustion.

The chemical properties of the alkanes

Because there are only C-H bonds throughout the whole structure of alkanes and these bonds are quite stable, and do not want to break easily, the alkanes are quite inert.

Alkanes undergo only two reactions:

1. Combustion reaction with oxygen: alkane + oxygen \Rightarrow carbon dioxide + water

For example: $CH_{4\,(g)} + \, 2O_{2\,(g)} \rightarrow CO_{2\,(g)} {+} 2H_2O_{(l)}$

2. Substitution reaction with halogens: alkane + halogen \rightarrow halogenoalkane + hydrogen halide

These types of reactions need sunlight or a UV light. For example:

 $\mathrm{CH}_{4\,\mathrm{(g)}}\mathrm{+Br}_{2\,\mathrm{(g)}}
ightarrow \mathrm{CH}_{3}\mathrm{Br}_{\mathrm{(g)}}\mathrm{+HBr}_{\mathrm{(g)}}$

The chemical properties of alkenes

In alkenes the functional group is a carbon-to-carbon double bond (C=C). This bond is quite unstable and breaks quite easily. This makes alkenes more reactive than alkanes. Unlike alkanes, alkenes undergo many reactions.

1. Combustion with oxygen: alkenes also react with oxygen in the air when it is ignited to produce carbon dioxide and water. But unlike alkanes they do not burn so cleanly.

For example: $C_2H_{4\,(g)}\!+\!O_{2\,(g)}\rightarrow 2CO_{2\,(g)}\!+\!2H_2O_{(l)}$

2. Addition reaction with halogens: the reaction normally takes place at room temperature and pressure. The reaction is called 'addition' because the two halogen atoms are being added to the hydrocarbon.

For example: ethene + bromine \rightarrow 1,2-dibromoethane

 ${\rm C_2H_4}_{\rm (g)} {+} {\rm Br_2}_{\rm (l)} \rightarrow {\rm C_2H_4Br_2}_{\rm (l)}$

3. Hydration: In the hydration reaction, ethene is reacted with steam to produce ethanol. The ethanol which is produced by the hydration process is extremely pure. Ethanol is produced in this way in factories.

For example: ethene + steam \rightarrow ethanol

 $C_2H_{4\,(g)}+\,H_2O_{(g)}\rightarrow C_2H_5OH_{(g)}$

The chemical properties of alcohols

The OH bond is quite polar which makes the alcohols quite reactive, so the alcohol undergoes several chemical reactions.

1. Reaction with oxygen: When ignited, alcohols burn or react with oxygen to produce carbon dioxide and water. Alcohol is a very good fuel. Alcohol produced by fermenting sugar cane (or other plant material) is used as a fuel is known as biofuel.

For example: ethanol + oxygen \rightarrow carbon dioxide + water

 ${
m C_2H_5OH_{(l)}\!+\!3O_{2(g)}}
ightarrow 2{
m CO}_{2(g)}\!+\!3{
m H_2O_{(l)}}$

2. Dehydration: Dehydration means the loss of water. In this reaction the alcohol loses a water molecule and forms an alkene of the same carbon number. So if ethanol undergoes the dehydration reaction, the corresponding alkene produced will be ethene.

The chemical properties of haloalkanes

Haloalkanes are quite reactive compounds. They undergo several different types of reactions. They also react with some metals to form organometallic compounds. The reactivity of haloalkanes is mainly due to the polar nature of the carbon-halogen bond.

Summary

In this unit you have learnt the following:

- The melting and boiling points of alkanes, alkenes and alkynes increase as the carbon number increases.
- Alkanes have relatively higher melting and boiling points than alkenes. This is because the presence of carbon-to-carbon double bond in alkenes reduces the effectiveness of the intermolecular forces between the molecules.
- Alkanes and alkenes are insoluble in water. This is because their weak London forces cannot disrupt the much stronger hydrogen bonds between the water molecules.
- Alcohols have much higher melting and boiling points compared to alkanes and alkenes as they can form hydrogen bonds.
- Alcohols are highly soluble in water as they can form hydrogen bonding with the water molecules.
- Alkanes do not have a functional group and undergo two chemical reactions: combustion and substitution.
- In alkenes the functional group is the carbon-to-carbon double bond. Alkenes undergo three chemical reactions: combustion, addition and hydration.
- In alcohols, the functional group is the OH group. Alcohols undergo two chemical reactions: combustion and dehydration.

Unit 3: Assessment

Suggested time to complete: 15 minutes

1. Which container (A or B) contains the compound with higher vapour pressure in it? Explain your answer.



2. Refer to the data table below which shows the melting points and boiling points for several organic compounds with different functional groups.

Formula	Name	Melting point (°C)	Boiling point (°C)
$ m C_4H_{10}$	butane	-137	0
$ m C_4H_8$	but-1-ene	-185	-6.5
$C_4H_{10}O$	butan-1-ol	-90	118
$\mathrm{C}_{5}\mathrm{H}_{12}$	pentane	-130	36
$\mathrm{C}_{5}\mathrm{H}_{10}$	pent-1-ene	-165.2	30
$C_5H_{12}O$	pentan-1-ol	-78	138

- a. At room temperature (approx. $25\,\,^\circ\mathrm{C}$), which of the organic compounds in the table are:
 - i. gases
 - ii. liquids
- b. Look at the alkane, alkene, and an alcohol with 4 carbon atoms:
 - i. How do their melting and boiling points compare?
 - ii. Explain why their melting points and boiling points are different?
- 3. The following diagram shows the reactants in an addition reaction:



- a. Draw the structural representation of the final product in this reaction.
- b. What is the condensed structural representation of the product?
- 4. Which types of homologous series can undergo additional reactions? Is this a saturated or unsaturated series?

The <u>full solutions</u> are at the end of the unit.

Unit 3: Solutions

Exercise 3.1

- 1. ethane and ethyne
- 2.
- A induced dipole
- B hydrogen bonding
- C hydrogen bonding
- 3. B has only induced-dipole forces (weak van der Waals forces). C has hydrogen bonding. Since hydrogen bonding is a stronger intermolecular force than van der Waals forces, more energy is required to separate the molecules of ethanol than the molecules of ethane. Thus ethanol has a higher boiling point than ethane.



Back to Exercise 3.1

Exercise 3.2

- 1. 1. butane and butene
- 2. The melting and boiling points increase as the number of carbon atoms increase. As the number of carbon atoms increases, so does the surface area of the molecule. This leads to stronger intermolecular forces, which are responsible for the increase in melting and boiling points.

Back to Exercise 3.2

Unit 3: Assessment

1. B. There are more molecules of B in the vapour phase than there are of A. More molecules in the vapour phase means a higher vapour pressure.

2.

- a.
- i. butane, but-1-ene
- ii. all except butane and but-1-ene

b.

- i. highest boiling (and melting) point is butan-1-ol, then butane (the alkane) followed by but-1-ene (the alkene).
- The alcohols can only form strong hydrogen bonds between molecules leading to higher melting and boiling points. The alkenes and alkanes form weak induced-dipole forces. However, the alkenes are unsaturated and so the melting and boiling points of an alkene are



b. $CH_3CH_2(Cl)$

4. Any unsaturated series where the functional group contains double or triple bonds (e.g. alkenes, alkynes).

Back to Unit 3: Assessment

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Unit 4: Applications of organic molecules

EMMA HARRAGE



What you should know

Before you start this unit, make sure you can:

- Understand hydrocarbons and functional groups. Refer to <u>level 4 subject outcome 5.3 unit 1</u> if you need help with this.
- Recall separation techniques. Refer to level 2 subject outcome 6.2 unit 2 if you need help with this.

Introduction

Parts of the text in this unit were sourced from <u>Siyavula Physical Science Gr 12 Learner's Book, Chapter 4</u>, released under a CC-BY licence.

In this unit you will learn about the formation of crude oil and how it is separated by fractional distillation into useful components.

Formation of crude oil

Fossil fuels are non-renewable sources of energy formed from the organic matter of plants and microorganisms that lived millions of years ago. This energy was originally captured via photosynthesis by living organisms such as plants, algae, and photosynthetic bacteria.

Fossil fuels are non-renewable because their formation took millions of years. Furthermore, higher productivity in the ancient environment allowed for more fossil fuel accumulation, meaning that the fossil fuel reserves available now could not necessarily be regenerated millions of years in the future.

Fossil fuels are composed primarily of hydrocarbons (molecules of just carbon and hydrogen), but they contain lesser amounts nitrogen, sulfur, oxygen, and other elements as well. Both oil and natural gas are fossil fuels found underground that formed from marine microorganisms. Oil (petroleum) is a liquid fossil fuel and consists of a variety of hydrocarbons while natural gas is a gaseous fossil fuel that consists of mostly methane and other small hydrocarbons.

Petroleum and natural gas formation

Tiny marine plants and animals died and were buried on the ocean floor. Over time, the marine plants and animals were covered by layers of silt and sand. Over millions of years, the remains were buried deeper and deeper. The enormous heat and pressure turned the remains into oil and natural gas. Today, we drill down through layers of sand, silt, and rock to reach the rock formations that contain oil and natural gas deposits.



Source: Adapted from National Energy Education Development Project (public domain)



Oil and natural gas formed from ancient marine microorganism (plankton). When plankton died, they were buried in sediments. As with coal, oxygen-poor conditions limited decomposition. As sediments continued to accumulate, the dead organisms were further buried. High temperature and pressure over millions of years ultimately produced oil and natural gas from these dead organisms.

Alkanes as fossil fuels

Crude oil when extracted is a thick black liquid. It is a complex mixture of different chemical compounds known as hydrocarbons ranging from methane to asphalt. The composition of crude oil varies with the type of crude oil and how it is extracted. Hydrocarbons such as alkanes are the main fraction of crude oil. It may also contain aromatic compounds and other compounds with nitrogen, and sulfur along with metals such as copper, nickel, vanadium, and iron. Hence crude oil is the main source of these organic compounds and fractional distillation is carried out to make these organic compounds available and usable.

Organic molecules vary in their physical and chemical properties due to their different carbon chain lengths and molecular masses. The melting point, boiling point, viscosity and colour are all related to the molecular mass and structure of the molecule. The boiling point of organic compounds increases with increasing molecular mass.

Fractional distillation

As mentioned, crude oil contains a wide range of organic compounds that differ in the length of their carbon chains and boiling points. Fractional distillation is the process to separate these individual fractions from each other. The whole process is based on the principle that different substances boil at different temperatures. For example, crude oil contains kerosene and naphtha, both are useful fractions but have different usage and applications. Naphtha can be used to form petrol for cars while kerosene is used for jet fuel. Both have different carbon chain lengths and structures. During distillation, the mixture of kerosene and naphtha is first vaporised then they are cooled down; the kerosene condenses at a higher temperature than the naphtha. As the mixture cools, the kerosene condenses first, and the naphtha condenses later. This is how fractional distillation works. The whole process can be divided into steps.

- The pre-treatment step: The crude oil extracted from an oil well could be an emulsion containing some rock, salt and water. In the first step, crude oil is pre-treated to remove this water and salt (sodium chloride or brine). The desalting step is important because the presence of sodium chloride can generate highly corrosive hydrogen chloride gas during the distillation process. This hydrogen chloride, being a strong acid may cause severe damage to the processing equipment. The desalting can be carried out by heating the crude oil under pressure or adding demulsifying agents e.g. soap, fatty acids, and surfactants to break the emulsion.
- Distillation: The crude oil is heated to 700 °C using heat exchangers and is converted to vapours. These vapours are sent to a distillation tower. The whole process is typically conducted in a vertical tower or column.

The tower has a temperature gradient, which means it is hot at the bottom and cooler at the top. It contains a series of trays installed to collect fractions that are finally removed via outlets. Because different fractions have different boiling points, they condense at different heights at a different level of the column. The large molecules with higher molecular mass stay in the bottom while smaller molecules with low molecular mass travel upward and condense at their respective collection tray. The solid waxy residue is removed from the bottom while molecules with very low molecular mass stay as gases and are collected from the top. They are called petroleum gas.



Figure 2: The distillation of crude oil

In figure 2, in diagram (a) you can see a distillation column used for separating petroleum fractions. In diagram (b) the petroleum fractions condense at different temperatures, depending on the number of carbon atoms in the molecules, and are drawn off from the column. The most volatile components (those with the lowest boiling points) condense at the top of the column, and the least volatile (those with the highest boiling points) condense at the bottom.

Hydrocarbon cracking is the process of breaking carbon-carbon bonds in long-chain hydrocarbons to form simpler, shorter-chain hydrocarbons.

There is a high demand for shorter alkanes but they are in short supply. Long chain hydrocarbons can be broken down into smaller chain alkanes and alkenes by cracking. Cracking is a thermal decomposition reaction. The hydrocarbon is vapourised and passed over a hot catalyst and this splits the long chains apart.

Hydrocarbon cracking is an important industrial process. These compounds include shorter alkanes and alkenes. A few examples of cracking hydrocarbons are given in figure 3.



Figure 3: The cracking of various hydrocarbons to produce alkanes and alkenes

It is possible to separate the products of hydrocarbon cracking and obtain specific products from a crude oil mix through fractional distillation. When the crude oil evaporates the gas bubbles through a tray that is kept at a certain temperature. The alkanes and alkenes that condense at that temperature will then condense in the tray. For example, if the tray is kept at 170 °C the product will be paraffin oil.

A fractionating column has a series of these trays (see figure 2), each at a constant temperature. This means that many compounds can be separated from the crude oil mix.

Bitumen for tar roads is collected at the bottom of the fractionating column. These are all compounds with more than 70 carbon atoms. The temperature decreases as you move up the column. As the gases rise, compounds with different length carbon chains condense until only the chains with 1 - 4 carbon atoms are collected at the top of the column. These are used for liquid petroleum gas.

$C_3 \dots C_4$	Propane and butane can be liquefied at low pressures, and are used, for example, in a propane gas burner, or as propellants in aerosol sprays. Butane is used in cigarette lighters.
$C_5 \dots C_8$	The alkanes from pentane to octane are highly volatile liquids and good solvents for nonpolar substances. They are used as fuels in internal combustion engines.
$C_9 \dots C_{16}$	Alkanes from nonane to hexadecane are liquids of higher viscosity, being used in diesel and aviation fuel (kerosene). The higher melting points of these alkanes can cause problems at low temperatures and in polar regions, where the fuel becomes too viscous.
$\mathrm{C}_{17}\mathrm{C}_{35}$	Alkanes with 17 to 35 carbon atoms form the major components of lubricating oil. They also act as anti-corrosive agents, as their hydrophobic nature protects the metal surface from contact with water. Solid alkanes are also used as paraffin wax in candles.
$> \mathrm{C}_{35}$	Alkanes with a chain length above 35 carbon atoms are found in bitumen (used in road surfacing). These higher alkanes have little chemical and commercial value and are usually split into lower alkanes by cracking.

N	ot	е

To consolidate your understanding of fractional distillation you can watch this video called Fractional Distillation by Fuse Schools.

Fractional Distillation (Duration: 04.04)



Combustion of alkanes

Alkanes are our most important fossil fuels. The combustion (burning) of alkanes is highly exothermic.

Methane is used as a fuel for domestic heating, cooking, and heating water. Methane is important for electricity generation by burning it as a fuel in a gas turbine or steam generator. Compared to other hydrocarbon fuels, methane produces less carbon dioxide for each unit of heat released. In many cities, methane is piped into homes for domestic heating and cooking.

In the complete combustion reaction of alkanes, carbon dioxide and water are released along with energy. Fossil fuels are burnt for the energy they release. The general reaction for the combustion of an alkane as a fossil fuel is:

alkane + oxygen \rightarrow carbon dioxide + water + energy

For example, the combustion of propane:

 $C_3H_{8(g)} + 5O_{2(g)} \rightarrow 3CO_{2(g)} + 4H_2O_{(g)} + energy$

Propane is a flammable gas used in blowtorches for soldering.



There are 2 oxygen atoms on the left. There are 13 oxygen atoms on the right. Divide the number of O atoms on the right by 2 to get $\frac{13}{2}$, this is the number of O₂ molecules required on the left:

$$\mathrm{C_4H_{10(g)}+}\; rac{13}{2}\mathrm{O}_{2(g)} \;
ightarrow \; 4\mathrm{CO}_{2(g)} + \; 5\mathrm{H_2O}_{(g)}$$

This is acceptable but it is better for all numbers to be whole numbers.

Step 4: Make sure all numbers are whole numbers

There is $\frac{13}{2}$ in front of the O₂ while all other numbers are whole numbers. So multiply the entire equation by₂:

 $2C_4H_{10(g)}+\ 13O_{2(g)}\ \ \rightarrow\ 8CO_{2(g)}+\ 10H_2O_{(g)}$

The combustion of hydrocarbons is an exothermic process otherwise there would not be much point in burning them to produce energy for fuel and heat.

Gasoline, which distils at temperatures between 40 - 205 °C is used as fuel for internal combustion engines. Gas oil, which distils at temperatures between 275 - 400 °C is used in diesel engines and industrial furnaces.

Summary

In this unit you have learnt the following:

- Crude oil can be separated by fractional distillation into simpler hydrocarbons.
- Simple hydrocarbons can be used as fuels
- Alkanes, including methane, are important as fuels because of the amount of heat they produce during combustion.

Unit 4: Assessment

Suggested time to complete: 15 minutes

- 1. Describe the conditions required for the formation of crude oil.
- 2. Briefly describe the process of fractional distillation.
- 3. Explain the relationship between the number of carbon atoms and the boiling point of the hydrocarbon.
- 4. Is the combustion of alkanes exothermic or endothermic? What does that mean?
- 5. Balance the following complete combustion equation: $C_7H_{16(l)}+O_{2(g)} \rightarrow CO_{2(g)}+H_2O_{(g)}$
- 6. Petrol is in fact not pure C_8H_{18} but a mixture of various alkanes. The octane rating of petrol refers to the percentage of the petrol which is C_8H_{18} . For example, 93 octane fuel contains 93% C_8H_{18} and 7% other alkanes.

Write a balanced equation for the chemical reaction which takes place when C_8H_{18} burns in oxygen.

The full solutions are at the end of the unit.

Unit 4: Solutions

Unit 4: Assessment

- 1. Intense heat and pressure from hundreds of layers of dead plants and animals being compressed together over millions of years.
- 2. Heating the crude oil mixture to different temperature ranges causes different fractions to vaporise. These vapours can be collected by cooling and condensing them in a separate receiving tube.
- 3. The higher the number of carbon atoms, the higher the boiling point. This is because more carbon atoms means larger molecules and stronger attractions between molecules, needing more energy to overcome the forces between them.
- 4. It is exothermic. This means that energy is released during the combustion of an alkane.
- 5. $C_7H_{16(l)} + 11O_{2(g)} \rightarrow 7CO_{2(g)} + 8H_2O_{(g)}$
- 6. $2C_8H_{18(l)} + 25O_{2(g)} \rightarrow 16CO_{2(g)} + 18H_2O_{(g)}$

Back to Unit 4: Assessment

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Unit 5: Plastics

EMMA HARRAGE



By the end of this unit you will be able to:

- Understand how plastics and polymers are made and refer to related South African industry.
- Understand the difference between thermoplastics and thermosets.

What you should know

Before you start this unit, make sure you can:

• Understand hydrocarbons. Refer to level 4 subject outcome 5.3 unit 1 if you need help with this.

Introduction

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In this unit you will learn about plastics: how they are made, what they are used for and plastic pollution. Plastic is a very versatile material. It can be easily shaped, it is waterproof, and it can be coloured.

What is plastic?

Plastic is defined as a material that contains, as an essential ingredient, an organic substance of large molecular mass.

The word, plastic, was derived from the word 'plastikos' meaning 'to mould' in Greek. Hydrocarbons from fossil fuels act as building blocks for long polymer molecules which are used to make different types of plastics.

- Plastics are extremely useful because they are:
- strong and ductile
- poor conductors of heat and electricity
- easily moulded into different shapes and sizes
- resist corrosion
- resistant to many chemicals.

Many companies in South Africa manufacture plastic products including Sasol which is a worldwide supplier in multiple industries such as construction and infrastructure, packaging, household and consumer goods, plastics and polymers, agriculture, industrial products, textile, health and medical supplies.

However, as we learnt in <u>level 2 subject outcome 7.1 unit 2</u>, plastics are not biodegradable and the incorrect disposal of them can contribute to pollution.

Plastics and polymers

Polymers are large molecules (macromolecules) that are made up of many repeating structural units called monomers which have various functional groups. To put it more simply, a monomer is like a building block.



Figure 1: When lots of monomers are joined together by covalent bonds, they form a polymer

In an organic polymer, the monomers are joined by the carbon atoms of the polymer backbone or chain. A polymer can also be inorganic, in which case there may be atoms such as silicon in the place of carbon atoms. We will look solely at organic polymers.

Polymers are a specific group of macromolecules. A macromolecule is any compound with many atoms.

The key feature that makes a polymer different from other macromolecules is the repetition of identical or similar monomers in the polymer chain. Polymers contain chains of the same type of functional group and that functional group is dependent on the monomer used.

Plastics are a group of polymers that can be moulded during manufacture. They can be one polymer or a blend (mixture) of polymers and may contain other substances as well. These other substances can be inorganic (e.g. used for electronic packaging) or stabilising (e.g. used for increasing fire resistance).

Depending on physical properties, plastics are divided into two types: thermoplastics and thermosetting plastics.

Polymers are formed through a process called polymerisation, where monomers react together to form a polymer chain. Two of the types of polymerisation reactions are addition polymerisation and condensation polymerisation.

The formation of polythene from ethene is an example of polymerisation. Small ethene monomers join to each other by the breaking of the double bond allowing them to join up to form one long carbon chain of poly(ethene).

For example:

propene → poly(propene)

Where propene is a monomer and polypropene is a polymer.

Addition polymerisation: In this type of reaction, monomer molecules are added to a growing polymer chain one at a time.

Condensation polymerisation: In this type of reaction, two monomer molecules combine by means of a covalent bond, and a small molecule such as water is lost in the bonding process. Nearly all biological macromolecules are formed using this process. Polyesters are polymers that form through condensation polymerisation.

Thermoplastics

Thermoplastics can be melted down at high temperatures and can cool to gain back solid form, and are generally of high molecular mass. The polymer chains are held together via intermolecular forces. The intermolecular forces can be easily broken by adding sufficient energy. This explains why this polymer is mouldable and will melt upon heating. When enough energy is provided to overcome the intermolecular forces that hold the polymer together as a solid, the solid will melt. When it cools, it gives off heat and the intermolecular forces re-form, making it a solid. Therefore, the process is reversible.

Examples of thermoplastic materials are polyethylene, polypropylene, polyvinyl chloride, polystyrene, polyamides, polyesters, and polyurethanes.

- Acrylonitrile butadiene styrene (ABS) is a thermoplastic used to manufacture:
 - sports equipment
 - toys
 - various automobile parts.
- Polycarbonate is used to make:
 - $_{\circ}~$ CDs and DVDs
 - drinking bottles
 - food storage containers
 - eyeglass lenses.
- Polyethylene is probably the most common thermoplastic and is used to make:
 - shampoo bottles
 - plastic grocery bags
 - bullet-proof vests.



Figure 2: Teflon tape

- The advantages of thermoplastics:
 - the primary advantage of thermoplastics is its wide range of applications
 - they are extremely adhesive to metal
 - thermoplastics are lightweight materials
 - they have high strength and have relatively low processing costs
 - they are highly recyclable
 - they have superb impact resistance and can be remoulded and reshaped
 - they have excellent corrosion resistance and detergent and chemical resistance
 - they provide electrical insulation.
- The disadvantages of thermoplastics:
 - thermoplastics degrade easily in direct sunlight or under UV exposure
 - not all thermoplastics resist hydrocarbons, organic solvents, and polar solvents
 - they can fracture rather than deform under high stress.

Thermosetting plastics

Thermosets are types of plastics that can withstand high temperatures without melting. This toughens and/ or hardens soft and viscous pre-polymers through the introduction of cross-links between polymer chains.

This process is known as 'curing' and it is done by heating the material above 200 °C and using additives. The cross-links are stable chemical bonds. Once the polymer is cross-linked, it becomes a very rigid and strong 3-D structure and will not melt upon heating.



Figure 3: Cross-linking makes the polymer much stronger

During the process of cross-linking, the molecular mass of the polymer increases; hence the melting point increases. When we heat up thermosets to uncontrollably high temperatures, they decompose instead of melting due to reaching the decomposition point before the melting point.

Thermosetting plastics are generally stronger than thermoplastic materials due to the cross-linking and are also better suited to high-temperature usage because they keep their shape as the strong covalent bonds between polymer chains cannot be broken easily.

These are some of the common uses of thermosetting polymers:

- \cdot to make kettles, plugs, laptop chargers, as they are heat resistant
- they are used in electrical fittings, handles and control knobs, and adhesives
- thermoset components are used extensively in a wide range of industries and are used for applications in the automotive, appliance, electrical, lighting, and energy markets due to excellent chemical and thermal stability along with superior strength, hardness, and moldability
- \cdot they are used to produce construction equipment panels
- they are used in agricultural equipment including feeding troughs and motors.

Some common examples of thermosets include polyester, fibreglass, polyurethanes, vulcanised rubber, acrylonitrile butadiene styrene (ABS), and melamine.



Figure 4: ABS is used to make Lego blocks

- The advantages of thermosetting plastics and their benefits over their metal counterparts:
 - the choice of colour and surface finishes
 - their resistance to corrosion effects and water
 - their low thermal conductivity and microwave transparency
 - their high strength-to-weight ratio and performance
 - they are cost-effective
 - their reduced production costs over fabrication using metals.
- There are certain disadvantages of thermosetting plastics :
 - they cannot be recycled
 - their low tensile strength and ductility
 - they cannot be remoulded or reshaped
 - their poor thermal conductivity when used for housing replacements
 - they are much more difficult to surface finish
 - the rigidity of the material can result in product failure when used in high vibration applications.

The difference between thermoplastic and thermoset plastic

The primary difference between the two is that thermoset is a material that strengthens when heated but cannot be remoulded or heated after the initial forming, while thermoplastics can be reheated, remoulded, and cooled as necessary without causing any chemical changes.

Thermoplastic	Thermoset plastic	
can be synthesised by the process called addition polymerisation	is synthesised by condensation polymerisation	
processed by injection moulding, extrusion process, blow moulding, thermoforming process, and rotational moulding	is processed by compression moulding, and reaction injection moulding	
has secondary bonds between molecular chains	has primary bonds between molecular chains and is held together by strong cross-links	
has a low melting point and low tensile strength	has a high melting point and tensile strength	
has lower molecular mass, compared to thermosetting plastic	has a high molecular mass	

Note

To consolidate your understanding of thermoset and thermoplastics, you can watch an edited version of this video, Fantastic Plastics. Or you can watch the full video on YouTube.

Fantastic Plastics - Edited (Duration: 03.50)



Fantastic Plastics – Full Version (Duration: 05.17)



Types of plastics

There are various types of plastic that are commonly used for many day-to-day applications.

Polyethylene

When lots of ethene molecules bond together, a polymer called polyethene (commonly called polyethylene) is formed. Ethene is the unsaturated monomer which, when joined to other ethene molecules through an addition reaction, forms the saturated polymer polyethene. Polyethene is the most common plastic with over 80 million metric tons produced each year. It is cheap and is used to make squeeze bottles, plastic bags, films, toys, and moulded objects as well as electric insulation.



Figure 5: Plastic bags are made from polyethenelt has a recycling number 4 which means that it is easy to process, has strength, toughness, flexibility, is easy to seal and is a barrier to moisture.

Polypropene (polypropylene)

Polypropene (commonly known as polypropylene) is stronger than polyethene and is used to make crates, fibres and ropes as well as being used in textiles, stationery, and car parts. In this polymer, the monomer is the alkene called propene.



Figure 6: Polypropene fabrics

Polyvinyl chloride (PVC)

Polyvinyl chloride or PVC is formed from the monomer chloroethene, which is commonly known as vinyl chloride. PVC is used in construction, especially to make plastic piping. With the addition of a plasticiser it is also used in clothing and upholstery and to replace rubber. The role of the plasticiser is to increase the ability of a material to change shape without breaking.



Figure 7: PVC tubing

Polystyrene

Polystyrene is made from the monomer styrene which is a liquid petrochemical. Polystyrene has many uses including protective packaging, trays, plastic lids and bottles.



Figure 8: Polystyrene packaging

Polyethylene terephthalate (PET)

Although there are many forms of polyesters, the term polyester usually refers to polyethylene terephthalate (PET). PET is made from an alcohol and an acid.

Polyesters have several characteristics which make them very useful. They are resistant to stretching and shrinking, they are easily washed and dry quickly, and they are resistant to mildew. It is for these reasons that polyesters are being used more and more in textiles. Polyesters are stretched out into fibres and can then be made into fabric and articles of clothing. In the home, polyesters are used to make clothing, carpets, curtains, sheets, pillows, and upholstery.





Figure 9: PET is used to make many different objects

Polyethylene terephthalate (PET) is not just a textile, it is also used to make plastic drink bottles. Many drink bottles are recycled by being reheated and turned into polyester fibres. This type of recycling helps to reduce disposal problems.

Acrylonitrile butadiene styrene (ABS)

Acrylonitrile butadiene styrene (ABS) is made from the polymerisation and emulsion of acrylonitrile, butadiene, and styrene. It has the following properties that make it ideal for industrial environments because it is resistant to corrosion:

- a high degree of rigidity
- good impact resistance at a wide range of temperatures
- good insulation, weldability, and strain restraint
- a high degree of stability and strength.

Most importantly, ABS polymers are chemically nonreactive, making them resistant to acids, alkalis, alcohols, and oils.

ABS thermoplastics are also a great substitute for metals in automotive manufacturing to reduce weight. Commonly used plastic car parts include dashboard components, seat backs, seat belt parts, door loners, handles, instrument panels and pillar trims.

Other common uses of ABS plastics include:

- home appliances and consumer goods such as control panels, housings for vacuum cleaners, food processors, refrigerator liners
- electrical and electronic applications such as computer keyboards and electronic enclosures
- construction applications such as pipes and fittings; this is because of it has high impact strength, and good resistance to rust and corrosion
- musical instruments
- sports equipment
- gardening tools
- Lego blocks and other plastic toys

• medical applications such as nebulisers.

Polylactic acid (PLA)

Despite the name, polylactic acid (PLA) is a polyester. It is an interesting polymer because the monomer used for this polymer comes from the biological fermentation of plant materials, while most monomers used in plastics come from petroleum. As a result PLA is biodegradable and has low carbon dioxide emissions. PLA is mostly used for packaging material and because it is biodegradable, it has the potential to alleviate land-fill disposal problems.

Plastics and the environment

Although plastics have had a huge impact globally, there is also an environmental price that must be paid for their use. The following are just some of the ways in which plastics can cause damage to the environment.

- 1. Waste disposal. Plastics are not easily broken down by micro-organisms and therefore most are not easily biodegradable. This leads to waste disposal problems.
- 2. Air pollution. When plastics burn, they can produce toxic gases such as carbon monoxide, hydrogen cyanide and hydrogen chloride. When polystyrene is burnt carbon monoxide is released. When PVC is burnt, hydrogen chloride and carbon monoxide are released.
- 3. Recycling. It is difficult to recycle plastics because each type of plastic has different properties and so different recycling methods may be needed for each plastic. Some plastics can be remelted and reused, while others can be ground up and used as a filler. One of the problems with recycling plastics is that they must be sorted according to plastic type.



Figure 10: A list of some of the different types and their identification codes; the alternative is that plastics should be re-used

In many countries, including South Africa, shoppers must now pay for plastic bags. This encourages people to collect and re-use the bags they already have.

Summary

In this unit you have learnt the following:

- A polymer is a macromolecule that is made up of many repeating structural units called monomers, which are joined by covalent bonds.
- Polymers that contain carbon atoms in the main chain are called organic polymers.
- Polymers form through a process called polymerisation. Two examples of polymerisation reactions are addition reactions and condensation reactions.
- An addition polymerisation reaction occurs when unsaturated monomers (e.g. alkenes) are added to each other one by one. The breaking of a double bond between carbon atoms in the monomer means that a bond can form with the next monomer.
- Thermoplastic is a plastic polymer material that becomes pliable or mouldable at a certain elevated temperature and solidifies upon cooling.
- Thermosetting plastic is a polymer that irreversibly becomes rigid when heated.
- Polyethene, polypropene, polyvinyl chloride, polyvinyl acetate and polystyrene are all polymers formed through addition polymerisation.
- Polyethylene terephthalate and polylactic acid are polymers formed through condensation polymerisation.
- Plastics are a subgroup of organic polymers that can be moulded. They can contain more than one polymer.
- It is not easy to recycle all plastics and so they create environmental problems.
- Some of these environmental problems include issues of waste disposal, air pollution and recycling.
- Polylactic acid is an example of a biodegradable polymer made from renewable resources.

Unit 5: Assessment

Suggested time to complete: 15 minutes

For questions 1-9, answer by giving the letter of the correct answer only.

- 1. What is the name of the small individual units that can be used to make plastics?
 - a. Alkanes
 - b. Polymers
 - c. Monomers
- 2. What is the main difference between thermoset and thermoplastic?
 - a. They can both be recycled.
 - b. Thermoset plastic cannot be heated and remoulded.
 - c. Thermoplastic cannot be heated and remoulded.
- 3. Which type of reaction occurs when plastics are made?
 - a. Fractional distillation
 - b. Cracking
 - c. Polymerisation
- 4. What does biodegradable mean?

- a. Something which is disposed of by burning.
- b. Can be broken down by bacteria in the soil and rots away.
- c. Something which can be easily reshaped.
- 5. What type of plastic is used in electrical plugs and sockets?
 - a. Biodegradable plastic
 - b. Thermoplastic
 - c. Thermoset plastics
- 6. Which type of material would be best suited for a purpose where low weight, low density and high strength and heat resistance are important?
 - a. Alloys
 - b. Ceramics
 - c. Plastics
- 7. What is the monomer used to make polythene?
 - a. Ethane
 - b. Ethene
 - c. Polythene
- 8. What is the polymer made from styrene?
 - a. Poly(ethene)
 - b. Poly(styrene)
 - c. Poly(styrane)
- 9. Which of these toxic gases can be released by burning polyvinyl chloride?
 - a. Hydrogen chloride (HCl)
 - b. Carbon monoxide (CO)
 - c. Hydrogen cyanide (HCN)
- 10. What is the main difference between thermoplastic and thermosetting plastic?
- 11. What is the main difference between the reactants used in an addition polymerisation and those used in a condensation polymerisation?

The <u>full solutions</u> are at the end of the unit.

Unit 5: Solutions

Unit 5: Assessment

- 1. c
- 2. b
- 3. c
- 4. b
- 5. c
- 6. c
- 7. b
- 8. b

- 9. c
- 10. The primary difference between the two is that thermoset is a material that strengthens when heated but cannot be remoulded or heated after the initial forming, while thermoplastics can be reheated, remoulded, and cooled as necessary without causing any chemical changes.
- 11. A condensation reaction requires two different monomer molecules, and an addition reaction requires only one kind of monomer.

Back to Unit 5: Assessment

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SUBJECT OUTCOME XI CHEMICAL CHANGE: IDENTIFY AND APPLY KNOWLEDGE OF ACIDS AND BASES



Subject outcome

Subject outcome 6.1: Identify and apply knowledge of acids and bases



Learning outcomes

- Define acid, base, and identify buffers using definitions of Arrhenius and Lowry Bronsted.
- Identify and apply pH values.
- Write neutralisation (acid-base) reactions down and calculate unknown value.
- Identify the use of indicators and titration.
- Identify industrial acids and alkali and their application.



Unit 1 outcomes

By the end of this unit you will be able to:

- Identify acids, bases and buffers using definitions of Arrhenius and Lowry Bronsted.
- Apply pH values.
- Write neutralisation (acid-base) reactions down and calculate unknown value.
- Explain the use of indicators and titration.
- Identify industrial acids and alkali and their application.

Unit 1: Acids and Bases

EMMA HARRAGE



Unit outcomes

By the end of this unit you will be able to:

- · Identify acids, bases and buffers using definitions of Arrhenius and Lowry Bronsted.
- Apply pH values.
- Write neutralisation (acid-base) reactions down and calculate unknown value.
- Explain the use of indicators and titration.
- · Identify industrial acids and alkali and their application.

What you should know

Before you start this unit, make sure you can:

- Balance chemical equations. Refer to level 3 subject outcome 6.2 unit 1 if you need help with this.
- Understand that there are different types of chemical reactions. Refer to <u>level 3 subject outcome 6.3</u> <u>unit 1</u> if you need help with this.

Introduction

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In this unit you will learn about acids and bases. An acid is defined as a substance which has H_3O^+ ions and the strength of an acid depends on how many of those ions it has. Strong acids include hydrochloric, nitric, and sulphuric acids which are used in laboratories and industry. Common acids include lemon juice (citric acid), vinegar (acetic acid, also called ethanoic acid), carbonic acid and lactic acid which is in milk. Bases are the chemical opposites of acids. An example of a strong bases is sodium hydroxide, and ammonia is a weak base. Household bases include baking soda, bicarbonate of soda, milk of magnesia, indigestion remedies, borax and most cleaning products including toothpaste and soaps. The pH scale gives a measure of the acidity or alkalinity of a solution and is determined using indicators.

What are acids and bases?

We encounter many examples of acids and bases in our daily lives. Some common examples of household items that contain acids are vinegar (contains acetic acid, also called ethanoic acid), lemon juice (contains citric and ascorbic acid), wine (contains tartaric acid). These acids are often found to have a sour taste. Hydrochloric acid, sulfuric acid and nitric acid are examples of acids that are more likely to be found in laboratories and industry.

Hydrochloric acid is also found in the gastric juices in the stomach. Fizzy (carbonated) drinks contain carbonic acid, while tea and wine contain tannic acid. People even use acids in an artistic process known as

acid etching. In acid etching, a metal is covered in a waxy material that is resistant to acid. The bare metal is then exposed in the desired pattern and the sample is placed in an acid bath.

Bases that you may know about include sodium hydroxide (commonly known as caustic soda), ammonium hydroxide and ammonia. Some of these are found in household cleaning products. Bases are usually found to have a bitter taste and feel slippery (soap is a good example). Acids and bases are also important commercial components in the fertiliser, plastics, and petroleum-refining industries.

Acid	Formula	Base	Formula
Hydrochloric acid	HCl	Sodium hydroxide	NaOH
Sulfuric acid	H_2SO_4	Potassium hydroxide	КОН
Nitric acid	HNO ₃	Magnesium hydroxide	$Mg(OH)_2$
Oxalic acid	$\mathrm{H}_{2}\mathrm{C}_{2}\mathrm{O}_{4}$	Calcium hydroxide	$Ca(OH)_2$
Sulfurous acid	H_2SO_3	Sodium bicarbonate	$NaHCO_3$
Phosphoric acid	H_3PO_4	Sodium carbonate	$\mathrm{Na}_2\mathrm{CO}_3$
Acetic (ethanoic) acid	CH ₃ COOH	Ammonium hydroxide	NH ₄ OH
Carbonic acid	H_2CO_3	Ammonia	$ m NH_3$

Table 1: Some common acids and bases and their chemical formulae

Models for acids and bases

For the acids you will encounter, an acid is a molecule that donates a ${\rm H}^+$ ion.

Substances that will act as a base include hydroxides, oxides, carbonates, or hydrogen carbonates, among others. Bases often release free hydroxide ions (OH^-) when dissociating in water.

Arrhenius model for acids and bases

A few models for acids and bases have been developed over the years. One of the earliest was the Arrhenius definition. In 1884, Arrhenius discovered that water dissociates (splits up) into hydronium ($_{\rm H_3O^+}$) and hydroxide ($_{\rm OH^-}$) ions according to the following equation:

$$2H_2O_{(1)} \rightleftharpoons H_3O^+{}_{(aq)}+OH^-{}_{(aq)}$$

Another way of writing this is:

$$H_2O_{(l)} \rightleftharpoons H^+{}_{(aq)} + OH^-{}_{(aq)}$$

Arrhenius described an acid as a compound that forms H_3O^+ when added to water. An Arrhenius acid therefore increases the concentration of H_3O^+ ions in water. Arrhenius described a base as a compound that dissociates in water to form OH^- ions. An Arrhenius base therefore increases the concentration of OH^- ions in water.


The definition of Arrhenius acids and bases: An Arrhenius acid forms H_3O^+ in water (increases H_3O^+). An Arrhenius base forms OH^- in water (increases OH^-).

Look at the following examples showing the dissociation of hydrochloric acid and sodium hydroxide: $HCl_{(aq)} + H_2O_{(l)} \rightarrow H_3O^+_{(aq)} + Cl^-_{(aq)}$ Hydrochloric acid in water increases the concentration of H_3O^+ ions and is therefore an acid.

Sodium hydroxide in water increases the concentration of OH $\,$ ions and is therefore a base: $\rm NaOH_{(s)}\,+\,H_2O_{(l)}\,\rightarrow\,Na^+{}_{(aq)}\,+\,OH_{(aq)}\,+\,H_2O_{(l)}$

However, this definition can only be used for acids and bases in water. Since there are many reactions which do not occur in water, it was important to come up with a much broader definition for acids and bases.

Brønsted-Lowry model for acids and bases

In 1923, Lowry and Brønsted took the work of Arrhenius further to develop a broader definition for acids and bases. The Brønsted-Lowry model defines acids and bases in terms of their ability to donate or accept protons.

Hydrogen atoms contain only one proton and one electron. ${
m H}^+$ is a hydrogen atom that has lost its electron and is often called a proton.



Under the Brønsted-Lowry definition an acid is a proton donor because it donates H^+ and a base is a proton acceptor because it accepts H^+ .

Below are some examples:

l. $\mathrm{HCl}_{(\mathrm{aq})} + \mathrm{NH}_{3(\mathrm{g})} \rightarrow \mathrm{NH}^{+4}_{(\mathrm{aq})} + \mathrm{CL}_{(\mathrm{aq})}$

To decide which substance is a proton donor and which is a proton acceptor, we need to look at what happens to each reactant. The reaction can be broken down as follows:

 $\mathrm{HCl}_{(\mathrm{aq})} \rightarrow \mathrm{H^{+}}_{(\mathrm{aq})} + \mathrm{Cl}_{(\mathrm{aq})}$

HCl donates a proton. It is a proton donor and is therefore the acid.

 $\mathrm{NH}_{3(\mathrm{g})}\,+\,\mathrm{H^+}_{(\mathrm{aq})}
ightarrow\mathrm{NH^{+4}}_{(\mathrm{aq})}$

 $\rm NH_3$ accepts a proton. It is a proton acceptor and is therefore the base.

2. $CH_3COOH_{(aq)} + H_2O_{(l)} \rightarrow H_3O^+_{(aq)} + CH_3COO^-_{(aq)}$

The reaction can be broken down as follows:

 $\mathrm{CH_3COOH_{(aq)}} ~\rightarrow~ \mathrm{CH_3COO^-_{(aq)}} ~+~ \mathrm{H^+_{(aq)}}$

 ${
m CH}_3{
m COOH}$ donates a proton. It is a proton donor and is therefore the acid.

 ${\rm H_2O_{(l)}\,+\,H^+{}_{(aq)}\rightarrow H_3O^+{}_{(aq)}}$

Water accepts a proton. It is a proton acceptor and is therefore the base.

3. $\rm NH_{3(g)} + H_2O_{(l)} \rightarrow \rm NH^{4+}{}_{(aq)} + OH^{-}{}_{(aq)}$

The reaction can be broken down as follows:

 $\mathrm{H_2O_{(l)}}~\rightarrow~\mathrm{OH^-_{(aq)}}+\mathrm{H^+_{(aq)}}$

Water donates a proton. It is a proton donor and is therefore the acid.

 $\rm NH_{3(g)}\,+\,{H^{+}}_{(aq)}\rightarrow \rm NH^{4+}{}_{(aq)}$

Ammonia accepts a proton. It is a proton acceptor and is therefore the base.

Notice that in example 2 water acted as a base, while in example 3 water acted as an acid. Water can act as both an acid and a base depending on the reaction. A substance that can act as either an acid or a base is called amphoteric or amphiprotic.

An **amphiprotic** substance can donate a proton in one reaction (a Brønsted-Lowry acid) or accept a proton in another reaction (a Brønsted-Lowry base).

Substances such as ammonia (NH_3), zinc oxide (ZnO), and beryllium hydroxide ($Be(OH)_2$) are amphoteric. Water and is also amphiprotic.

Weak acids and bases

A weak acid or base is one where only a small percentage of molecules will dissociate to form ions in solution.



Hydrogen fluoride (HF) is an example of a weak acid. For example if 100 000 molecules of HF are added to water and only 100 ionise to form H^+ and F^- ions, then there is only a small amount of ionisation. This is what makes HF a weak acid.

 $HF(g) + H_2O(I) \longrightarrow H_3O^+(aq) + F^-(aq)$

very little ionisation small percentage of molecules ionised

Figure 1: The unequal double arrows in the reaction equation indicate that the equilibrium position does not favour the formation of ions

An example of a weak base is magnesium hydroxide $(Mg(OH)_2)$, which will only dissociate partially into Mg^{2+} and OH^- ions.

Dilute and concentrated solutions

A different concept to strong and weak is the concept of concentrated and dilute. Where strong and weak refer to the characteristic of a compound, concentrated and dilute refer to the characteristic of a solution. Thus a strong acid can be prepared as either a concentrated solution or a dilute solution. A solution of which the exact concentration is known is called a standard solution.

- A standard solution is one where the exact concentration of solute in a solvent is known.
- A concentrated solution is one where a large amount of a substance (solute) has been added to a solvent. Note that both strong and weak acids and bases can be used in concentrated solutions.
- A dilute solution is one where a small amount of a substance has been added to a solvent. Note that both strong and weak acids and bases can be used in dilute solutions.

A concentrated solution can be made from a strong or a weak acid or base. A dilute solution can also be made from a strong or a weak acid or base. Whether a solution is concentrated or dilute depends on how much of the acid or base was added to the solvent.

A strong base that is also concentrated would be a base that almost completely dissociates when added to a solution, and you also add a large amount of the base to the solution.

A weak acid that is also dilute would be an acid where only a small percentage of molecules ionise when added to a solution, and you also add only a small amount of the acid to the solution.

Table 2: A summary of the properties of strong, weak, concentrated, and dilute acids and bases

	Acid	Base
Strong	high percentage forms ions in solution	high percentage forms ions in solution
Weak	only a small percentage forms ions in solution	only a small percentage forms ions in solution
Concentrated	large number of moles of acid in solution	large number of moles of base in solution
Dilute	small number of moles of acid in solution	small number of moles of base in solution

There are only six strong inorganic acids, the rest are considered weak. These are:

- HClO₄ (perchloric acid)
- HI(hydroiodic acid)
- HBr(hydrobromic acid)

- HCI(hydrochloric acid)
- \cdot H₂SO₄(sulfuric acid)
- HNO₃ (nitric acid).



Example 1.1

Solution 1 contains $_{100}~\rm dm^3$ of HCl added to $_{10}~\rm dm^3$ of water. Almost all the HCl molecules ionise in the solution.

Solution 2 contains 0.01 g of $Mg(OH)_2$ added to $1\ 000\ dm^3$ of water. Only a small percentage of the $Mg(OH)_2$ molecules dissociate in the solution.

Say whether these solutions:

- 1. Contain a strong or weak acid or base.
- 2. Are concentrated or dilute.
- 3. Are the compounds acids or bases?

Solution

HCl is hydrochloric acid. It would donate a proton and is an acid.

 $\mathrm{Mg(OH)}_2$ is magnesium hydroxide and is a base.

Step1: What makes an acid or base strong or weak?

Almost complete ionisation or dissociation means an acid or base is strong. Only a small amount of ionisation or dissociation means an acid or base is weak.

Step 2: What makes a solution concentrated or dilute?

A concentrated solution has a high ratio of solute to solvent. A dilute solution has a low ratio of solute to solvent.

Are the solutions concentrated or dilute?

 $100~dm^3$ of HCl is added to $10~dm^3$ of water. This is a high ratio, therefore the solution of HCl is concentrated.

 $0.01 \text{ g of } Mg(OH)_2$ is added to $1 \ 000 \ dm^3$ of water. This is a low ratio, therefore the solution of $Mg(OH)_2$ is dilute.

Step 3: Are the compounds strong or weak acids and bases?

Almost all the HCl molecules ionise in the solution, therefore HCl is a strong acid.

Only a small percentage of the $Mg(OH)_2$ molecules dissociate, therefore $Mg(OH)_2$ is a weak base.

Step 4: Combine your information

Solution 1 is a concentrated solution of a strong acid.

Solution 2 is a dilute solution of a weak base.

To calculate the concentration of a solution we use the formula: $c \pmod{\text{mol.dm}^{-3}} = \frac{n \pmod{1}}{V \pmod{3}}$



Example 1.2

 $0.27~{
m g}$ of ${
m H}_2{
m SO}_4$ is added to ${
m 183.7~dm^3}$ of water. Calculate the concentration of the solution.

Solution

Step 1: List the information you have and the information you need

 $V = 183.7 \text{ dm}^3$, m = 0.27 g

The volume (V) and the mass (m) are given. The number of moles (n) needs to be calculated. To do that the molar mass (M) needs to be calculated.

Step 2: Make sure all given units are correct and convert them if necessary

All the units are correct.

Step 3: What equations will be necessary to calculate the concentration?

$$egin{aligned} &c \left(\mathrm{mol.dm}^{-3}
ight) = rac{n \ (\mathrm{mol})}{V \ (\mathrm{dm}^3)} \ &n \ (\mathrm{mol}) = rac{m \ (\mathrm{g})}{M \ (\mathrm{g.mol}^{-1})} \end{aligned}$$

Step 4: Calculate the number of moles of the acid in the solution

$$egin{aligned} M_{(H_2SO_4)} &= (2 \, \mathrm{x} \, 1.01) + 32.1 + (4 \, \mathrm{x}16) = 98.12 \, \mathrm{g.mol^{-1}} \ n &= rac{m}{M} = rac{0.27}{98.12} = 0.0028 \, \mathrm{mol} \end{aligned}$$

Step 5: Calculate the concentration of the solution

$$c = rac{n}{V} = rac{0.0028 ext{ mol}}{183.7 ext{ dm}^3} = 0.0000152 ext{ mol.dm}^{-3} = 1.52 imes 10^{-5} ext{ mol.dm}^{-3}$$

Example 1.3

 $16.4 \mathrm{g}$ of KOH is added to $12.9 \mathrm{~cm^3}$ of water. Calculate the concentration of the solution.

Solution

 \bigcirc

Step 1: List the information you have and the information you need

 $V = 12.9 \text{ cm}^3$, m = 16.4g

The volume (V) and the mass (m) are given. The number of moles (n) needs to be calculated. To do that the molar mass (M) needs to be calculated.

Step 2: Make sure all given units are correct or convert them

The volume needs to be converted to dm^3 .

$$V ~=~ 12.9~{
m cm}^3 imes {0.001~{
m dm}^3\over 1~{
m cm}^3} = 0.0129~{
m dm}^3$$

Step 3: What equations will be necessary to calculate the concentration?

$$egin{aligned} c \left(\mathrm{mol.dm}^{-3}
ight) &= rac{n \ (\mathrm{mol})}{V \ (\mathrm{dm}^3)} \ n \ (\mathrm{mol}) &= rac{m \ (\mathrm{g})}{M \ (\mathrm{g.mol}^{-1})} \end{aligned}$$

Step 4: Calculate the number of moles of base in the solution

$$egin{aligned} M_{
m (KOH)} &= 39.1 + 16 + 1.01 = 56.11 \, {
m g.mol}^{-1} \ n &= rac{m}{M} = rac{16.4 \, {
m g}}{56.11 \, {
m g.mol}^{-1} = \ 0.292 \, {
m mol}} \end{aligned}$$

Step 5: Calculate the concentration of the solution

$$c = \frac{n}{V} = \frac{0.292 \text{ mol}}{0.0129 \text{ dm}^3} = 22.64 \text{ mol.dm}^{-3}$$



Exercise 1.1

- 1. Say whether the solutions of the acids and bases in the following situations are concentrated or dilute.
 - a. For every 1 mole of a solvent there are 50 moles of lithium hydroxide (LiOH).
 - b. For every 100 moles of a solvent there are 5 moles of nitric acid (HNO₃).
- 2. Calculate the concentration of the following solutions:
 - a. 27 g of sodium bicarbonate (NaHCO_3) added to $22.6~{
 m cm}^3$ of water.
 - b. 0.893 mol of phosphoric acid (H₃PO₄) added to 4.79 dm^3 of a solvent.

The <u>full solutions</u> are at the end of the unit.

Acid-base reactions

When an acid and a base react, they form a salt. If the base contains hydroxide ($_{OH^-}$) ions, then water will also be formed. The word salt is a general term which applies to the products of all acid-base reactions. A salt is a product that is made up of the cation from a base and the anion from an acid.

A cation is an ion (charged atom or molecule) with a positive (+) charge. An anion is an ion with a negative (-) charge.

Salts are not just the table salt you put on your food. A salt is any compound made up of stoichiometrically equivalent amounts of cations and anions to make a neutral, ionic compound.



Potassium permanganate

Copper sulfate

Potassium dichromate

Figure 2: Examples of salts

When an equivalent amount of acid and base react (so that neither the acid nor the base are in excess), the reaction is said to have reached the equivalence point. At this point neutralisation has been achieved.

Look at the following examples:

1. Hydrochloric acid with sodium hydroxide:

Hydrochloric acid reacts with sodium hydroxide to form sodium chloride (the salt) and water. Sodium chloride is made up of Na^+ cations from the base (NaOH) and Cl⁻ anions from the acid (HCl).

 $\mathrm{HCl}_{(\mathrm{aq})}\!+\!\mathrm{NaOH}_{(\mathrm{aq})}\rightarrow\mathrm{NaCl}_{(\mathrm{aq})}\!+\!\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}$

2. Hydrogen bromide with potassium hydroxide:

Hydrogen bromide reacts with potassium hydroxide to form potassium bromide (the salt) and water. Potassium bromide is made up of K^+ cations from the base (KOH) and Br^- anions from the acid (HBr).

 $\mathrm{HBr}_{(\mathrm{aq})}\!+\!\mathrm{KOH}_{(\mathrm{aq})}\rightarrow\mathrm{KBr}_{(\mathrm{aq})}\!+\!\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}$

3. Hydrochloric acid with sodium hydrogen carbonate:

Hydrochloric acid reacts with sodium hydrogen carbonate to form sodium chloride (the salt), water and carbon dioxide. Sodium chloride is made up of $_{Na^+}$ cations from the base (NaHCO₃) and Cl⁻ anions from the acid (HCl).

 $\mathrm{HCl}_{(\mathrm{aq})} \! + \! \mathrm{NaHCO}_{3(\mathrm{aq})} \rightarrow \mathrm{NaCl}_{(\mathrm{aq})} \! + \! \mathrm{H}_2\mathrm{O}_{(\mathrm{l})} \! + \! \mathrm{CO}_{2(\mathrm{g})}$

You should notice that in the first two examples, the base contained OH^- ions, and therefore the products were a salt and water. NaCl (table salt) and KBr are both salts. In the third example, $NaHCO_3$ also acts as a base, despite not having OH^- ions. A salt is still formed as one of the products, but carbon dioxide (CO_2) is produced as well as water.

Neutralisation reactions are very important in everyday life. Below are some examples:

- Domestic uses: Calcium oxide (CaO) is used to neutralise acidic soil. Powdered limestone (CaCO₃) can also be used, but its action is much slower and less effective. These substances can also be used on a larger scale in farming and in rivers.
- Biological uses: Hydrochloric acid (HCI) in the stomach plays an important role in helping to digest food. It is important to note that too much acid in the stomach may lead to the formation of ulcers in cases where the stomach lining is damaged (e.g. by an infection). Antacids (which are bases) are taken to neutralise excess stomach acid, to prevent damage to the intestines. Examples of antacids are aluminium hydroxide, magnesium hydroxide (milk of magnesia) and sodium bicarbonate (bicarbonate of soda).
- \cdot Industrial uses: Alkaline calcium hydroxide (limewater) is used to absorb harmful acidic SO_2 gas that is

released from power stations and from the burning of fossil fuels.





Determining equations from starting materials:

Magnesium carbonate ($MgCO_3$) is dissolved in nitric acid (HNO_3). Give the balanced chemical equation for this reaction.

Solution

Step 1: What are the reactants?

An acid (HNO_3) and a metal carbonate $(MgCO_3)$.

Step 2: What will the products be?

As this is the reaction of an acid and a metal carbonate the products will be a salt, water, and carbon dioxide.

nitric acid + magnesium carbonate \rightarrow salt + water + carbon dioxide

Step 3: What is the formula of the salt?

The cation will come from the metal carbonate (Mg^{2+}) . The anion will come from the acid (NO^{3-}) . Due to the charges on the cation and anion there must be two NO^{3-} for everyone Mg^{2+} .

Therefore the formula for the salt will be $Mg(NO_3)_2$.

Step 4: Write the equation for this reaction

$$HNO_{3(aq)} + MgCO_{3(s)} \ \rightarrow \ Mg(NO_3)_{2(aq)} + H_2O_{(l)} + CO_{2(g)}$$

Step 5: Make sure that the equation is balanced

The equation is not balanced.

	Number on left	Number on right
н	1	2
N	1	2
0	6	9
Mg	1	1
с	1	1

To balance this equation there needs to be two nitric acid molecules on the left-hand side.

 $2HNO_{3(aq)} + MgCO_{3(s)} \ \rightarrow \ Mg(NO_3)_{2(aq)} + H_2O_{(l)} + CO_{2(g)}$

	Number on left	Number on right
н	2	2
N	2	2
0	9	9
Mg	1	1
с	1	1

The equation is now balanced.

Example 1.5

Hydroiodic acid (HI) is added to solid potassium hydroxide (KOH). Give the balanced chemical equation for this reaction.

Solution

Step 1: What are the reactants?

An acid (HI) and a base (KOH).

Step 2: What will the products be?

As this is the reaction of an acid and base (which contains a hydroxide anion) the products will be a salt and water.

hydroiodic acid + potassium hydroxide → salt + water

Step 3: What is the formula of the salt?

The cation will come from the base (K^+). The anion will come from the acid (I^-). Due to the charges on the cation and anion there must be one K^+ for everyone I^- .

Therefore the formula for the salt will be ${
m KI}$.

Step 4: Write the equation for this reaction

 $\mathrm{HI}_{(\mathrm{aq})}\!+\!\mathrm{KOH}_{(\mathrm{s})}\rightarrow\mathrm{KI}_{(\mathrm{aq})}\!+\!\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}$

This can also be written: $HI_{(aq)} + KOH_{(s)} \rightarrow K^{+}{}_{(aq)} + I^{-}{}_{(aq)} + H_2O_{(l)}$

Step 5: Make sure that the equation is balanced

The equation is balanced.

₁₂O

Example 1.6

Sulfuric acid (H_2SO_4) and ammonia (NH_3) are combined. Give the balanced chemical equation for this reaction.

Solution

Step 1: What are the reactants?

An acid (H_2SO_4) and a base (NH_3) .

Step 2: What will the products be?

As this is the reaction of an acid and a base (with no hydroxide anion), there will be a salt as a product. There may or may not be another product.

sulfuric acid + ammonia → salt (+ maybe another product)

Step 3: What is the formula of the salt?

The cation will come from the base accepting an H^+ (NH^{+4}). The anion will come from the acid (SO_2^{-4}). Due to the charges on the cation and anion there must be two NH^{4+} for everyone SO_2^{-4-} .

Therefore the formula for the salt will be $(NH_4)_2SO_4$.

Step 4: Write the equation for this reaction so far

 $H_2SO_{4(aq)}+\,NH_{3(g)}\ \rightarrow (NH_4)_2SO_{4(aq)}\ (+\ maybe\ another\ product)$

Step 5: Determine if there will be another product

There are no atom types that are not accounted for on both sides of the equation, therefore it is unlikely that there will be another product. If the equation can be balanced, then there is no other product.

Step 6: Make sure that the equation is balanced

The equation is not balanced.

	Number on left	Number on right
н	5	8
S	1	1
0	4	4
N	1	2

To balance this equation there needs to be two ammonia molecules on the left-hand side.

 $\mathrm{H}_2\mathrm{SO}_{4(\mathrm{aq})} + \ 2\mathrm{NH}_{3(\mathrm{g})} \ \rightarrow (\mathrm{NH}_4)_2\mathrm{SO}_{4(\mathrm{aq})}$

	Number on left	Number on right
н	8	8
s	1	1
0	4	4
N	2	2

The equation is now balanced.



Exercise 1.2

Write balanced equations for these reactions:

- 1. Hydrochloric acid and calcium
- 2. Hydrochloric acid and magnesium hydroxide
- 3. Sulfuric acid and magnesium oxide

The <u>full solutions</u> are at the end of the unit.

The pH scale

The concentration of specific ions in solution determines whether the solution is acidic or basic. Acids and bases can be described as substances that either increase or decrease the concentration of hydrogen ($_{\rm H^+}$) or hydronium ($_{\rm H_3O^+}$) ions in a solution. An acid increases the hydrogen ion concentration in a solution, while a base decreases the hydrogen ion concentration.

pH is used to measure the concentration of H^+ ions ([H^+]) and therefore, whether a substance is acidic or basic **(alkaline)**. Solutions with a pH of less than seven are acidic, while those with a pH greater than seven are basic (alkaline). The pH scale ranges from 0 to 14 and a pH of 7 is considered neutral.



Figure 3: The universal indicator changes colour from red in strongly acidic solutions through to purple in strongly basic solutions



The term pH was first used by in 1909 by Søren Peter Lauritz Sørensen (a Danish biochemist). The p stood for potenz and the H for hydrogen. This translates to power of hydrogen.

Table 3: The pH of solutions of acids and bases as found in common household items

Molecule	Found in	рН	Туре
Phosphoric acid	Fizzy drinks	2.15	acid
Tartaric acid	Wine	2.95	acid
Citric acid	Lemon juice	3.14	acid
Acetic acid	Vinegar	4.76	acid
Carbonic acid	Fizzy drinks	6.37	acid
Ammonia	Cleaning products	11.5	base
Ammonium hydroxide	Cleaning products	11.65	base
Sodium hydroxide	Caustic soda	13	base

A strong acid + strong base form a neutral salt and water solution \rightarrow pH =7

For example: $H_2SO_{4(l)} + NaOH_{(s)} \rightarrow Na_2SO_{4(aq)} + H_2O_{(l)}$

A weak acid + strong base form a weak basic salt and water solution → pH = approximately 9

For example: $HF_{(l)}\!+\!NaOH_{(s)}\rightarrow NaF_{(aq)}\!+\!H_2O_{(l)}$

A strong acid + weak base form a weak acidic salt and water solution → pH = approximately 5

 $\rm H_2SO_{4(l)} + NH_{3(l)} \rightarrow (NH_4)_2SO_{4(aq)} + H_2O_{(l)}$

In agriculture, it is important for farmers to know the pH of their soils so that they can plant the right kinds of crops. The pH of soils can vary depending on several factors, such as rainwater, the kinds of rocks and materials from which the soil was formed and human influences such as pollution and fertilisers. The pH of rainwater can also vary, and this too influences agriculture, buildings, water courses, animals, and plants. Rainwater is naturally acidic because carbon dioxide in the atmosphere combines with water to form carbonic acid. Unpolluted rainwater has a pH of approximately 4.5. However, human activities can alter the acidity of rain, and this can cause serious problems such as acid rain.

Indicators

Indicators are chemical compounds that change colour depending on whether they are in an acidic or a basic solution. Different indicators are used to identify acids and bases. Only a universal indicator can identify an acid and base and give a pH value.



Figure 5: Different types of indicators

Litmus paper can be used as an acid/base indicator. It is sold in strips. Blue litmus paper will become red in acidic conditions and red litmus paper will become blue in basic conditions. Blue litmus paper is used to detect acidic conditions, while red litmus paper is used to detect basic conditions.



Figure 6: Blue litmus indicates neutral and acidic, red litmus indicates neutral and basic

	Preferred indicator	Colour of acid	Colour of base	pH range
strong acid + strong base	bromothymol blue	yellow	blue	6.0-7
weak acid + strong base	phenolphthalein	colourless	pink	8.3 - 10
strong acid + weak base	bromocresol green	yellow	blue	3.8 - 5.4

Table 4: Different indicators with their pH ranges

Titrations

The neutralisation reaction between an acid and a base can be very useful. If an acidic solution of known concentration (a standard solution) is added to a basic (alkaline) solution of unknown concentration until the solution is exactly neutralised (i.e. there is only salt and water), it is possible to calculate the exact concentration of the unknown solution. It is possible to do this because, at the exact point where the solution is neutralised, stoichiometrically equivalent mole amounts of acid and base have reacted with each other.

In a titration, a known volume of a standard solution (A) is added to a known volume of a solution with unknown concentration (B). The concentration of B can then be determined.

Acids and bases are commonly used in titrations, and the point of neutralisation is called the endpoint of the reaction. If you have an indicator that changes colour in the range of the endpoint pH, then you will be able to see when the endpoint has occurred. Another name for a titration is volumetric analysis.



Figure 8: Equipment used in a titration experiment

Note

To see an actual titration, watch this video by Fuse Schools, called How to do titrations.

How to do titrations (Duration: 03.42)



Titration calculations

So how exactly can a titration be carried out to determine an unknown concentration? Look at the following steps to help you to understand the process.

- 1. A carefully measured volume of the solution with unknown concentration is put into a conical flask.
- 2. A few drops of a suitable indicator is added to this solution (bromothymol blue and phenolphthalein are common indicators, refer to table 4).
- 3. The conical flask is placed on a white tile or piece of paper (to make colour changes easier to see).
- 4. A volume of the standard solution (known concentration) is put into a burette (a measuring device) and is slowly added to the solution in the flask, drop by drop.
- 5. At some point, adding one more drop will change the colour of the unknown solution to the colour of the endpoint of the reaction.
- 6. Record the volume of standard solution that has been added up to this point.
- 7. Use the information you have gathered to calculate the exact concentration of the unknown solution. Worked examples are given to walk you through this step.
- 8. Note that adding more solution once the endpoint has been reached will result in a colour change from the endpoint colour to that of the acid (if the solution in the conical flask is a base) or of the base (if the solution in the conical flask is an acid).

When you are busy with these calculations, you will need to remember the following:

- $\cdot~1~dm^3=1~L=1~000~mL=1~000~cm^3$, therefore dividing cm^3 by 1 000 will give you an answer in dm^3
- $\cdot~$ the concentration of a solution is measured in $\rm mol.dm^{-3}$
- \cdot $\,$ make sure all the units are correct in your calculations.

Example 1.7

Given the equation: $NaOH_{(aq)} + HCl_{(aq)} \ \rightarrow NaCl_{(aq)} + H_2O_{(l)}$

 25 cm^3 of sodium hydroxide solution was pipetted into a conical flask and titrated with 0.2 mol.dm^{-3} hydrochloric acid. Using a suitable indicator, it was found that 15 cm^3 of acid was needed to neutralise the base. Calculate the concentration of the sodium hydroxide.

Solution

Step 1: Make sure that the equation is balanced

There are equal numbers of each type of atom on each side of the equation, so the equation is balanced.

Step 2: Write down all the information you know about the reaction, converting to the correct units

NaOH:
$$V = 25 \text{ cm}^3 \times \frac{0.001 \text{ dm}^3}{1 \text{ cm}^3} = 0.025 \text{ dm}^3$$

HCl: $V = 15 \text{ cm}^3 \times \frac{0.001 \text{ dm}^3}{1 \text{ cm}^3} = 0.015 \text{ dm}^3$ $C = 0.2 \text{ mol.dm}^{-3}$

Step 3: Calculate the number of moles of HCl that are added

 $c = rac{n}{V}$

Therefore, $n_{(\rm HCl)} = c \times V$ $n_{(HCl)} = 0.2 \, \, {\rm mol.dm^{-3}} \, \times \, 0.015 \, \, {\rm dm^3} = 0.003 \, \, {\rm mol}$ $0.003 \, \, {\rm moles}$ of HCl are required to neutralise the base.

Step 4: Calculate the number of moles of sodium hydroxide in the reaction

Look at the equation for the reaction: the molar ratio of HCl:NaOH is 1:1.

So for every mole of $\rm HCl$, there is one mole of $\rm NaOH$ that is involved in the reaction. Therefore, if $0.003~\rm mol~HCl$ is required to neutralise the solution, $0.003~\rm mol~NaOH$ must have been present in the sample of the unknown solution.

Calculate the concentration of the sodium hydroxide solution:

$$c_{(NaOH)} = rac{n}{V} = rac{0.003}{0.025} = 0.12 \,\, {
m mol.dm^{-3}}$$

The concentration of the NaOH solution is 0.12 mol.dm^{-3}

_{Ja}O

Example 1.8

 $10~{\rm g}$ of solid sodium hydroxide is dissolved in $500~{\rm cm}^3$ water. Using titration, it was found that $20~{\rm cm}^3$ of this solution was able to completely neutralise $10~{\rm cm}^3$ of a sulfuric acid solution. Calculate the concentration of the sulfuric acid.

Solution

Step 1: Write a balanced equation for the titration reaction

The reactants are sodium hydroxide (NaOH) and sulfuric acid (H_2SO_4). The base has a hydroxide anion (OH-), therefore the products will be a salt and water.

The cation for the salt (Na^+) will come from the base. The anion for the salt (SO_2^{-4}) will come from the acid. There must be $2 Na^+$ cations for everyone SO_2^{-4} and the salt will be Na_2SO_4 .

To balance the equation we need to multiply the number of sodium hydroxide molecules and the water molecules by two.

 $\mathrm{H_2SO_4}_{(\mathrm{aq})} + 2\mathrm{NaOH}_{(\mathrm{aq})} \rightarrow \ \mathrm{Na_2SO_4}_{(\mathrm{aq})} + 2\mathrm{H_2O}_{(\mathrm{l})}$

Step 2: Calculate the concentration of the sodium hydroxide solution

The total volume that the 10 g was dissolved in must be used to calculate the concentration.

$$V = 500 \,\, {
m cm}^3 \ imes {0.001 \over 1} = \ 0.5 \,\, {
m dm}^3$$

 $M_{
m (NaOH)} = 23.0 ~+~ 16.0 ~+~ 1.01 = 40.01 {
m ~g.mol}^{-1}$

$$n_{(\mathrm{NaOH})} = rac{m}{M} = rac{10}{40.01} = 0.25 ext{ mol}
onumber \ C_{(\mathrm{NaOH})} = rac{n}{V} = rac{0.25}{0.5} = 0.50 ext{ mol.dm}^{-3}$$

Step 3: Calculate the number of moles of sodium hydroxide that were used in the neutralisation reaction

Remember that only $m _{20~cm^3}$ of the sodium hydroxide solution is used:

 $V = 20 \, {
m cm}^3 \ imes {0.001 \over 1} = 0.02 \, {
m dm}^3$ $c = {n \over V} \, {
m therefore} \, n = c imes V$ $n = 0.50 imes 0.02 = 0.01 \, {
m mol}$

Step 4: Calculate the number of moles of sulfuric acid that were neutralised

According to the balanced chemical equation, the mole ratio of N_aOH to H_2SO_4 is 2:1. There are 2 moles of N_aOH for every 1 mole of H_2SO_4 .

 $n_{
m (H_2SO_4)} = {0.01\over 2} = 0.005 \,\,{
m mol}$

Step 5: Calculate the concentration of the sulfuric acid solution

Remember that $10\,\,cm^3$ of the sulfuric acid solution is neutralised.

$$V_{
m (H_2SO_4)} 10 {
m cm}^3 imes {0.001 \over 1} = 0.01 {
m dm}^3 \ c = {n \over V} = {0.005 \over 0.01} = 0.5 {
m mol.dm}^{-3}$$

Note

To help you with titration calculations you can watch this video by Fuse Schools, called Titration calculations.

Titration calculations (Duration: 05.12)





Exercise 1.3

- 1. A learner is asked to prepare a standard solution of the weak acid, oxalic acid $(H_2C_2O_4)$, for use in a titration. The volume of the solution must be 500 cm^3 and the concentration must be 0.2 mol.dm^{-3} . Calculate the mass of oxalic acid which the learner must dissolve to make up the required standard solution.
- 2. The learner titrates this 0.2 mol.dm^{-3} oxalic acid solution against a solution of sodium hydroxide. He finds that 40 cm^3 of the oxalic acid solution completely neutralises 35 cm^3 of the sodium hydroxide solution.

Calculate the concentration of the sodium hydroxide solution.

The <u>full solutions</u> are at the end of the unit.

Applications of acids and bases

There are a variety of important uses of acids in bases.

The production of chlorine

The chlorine-alkali (chloralkali) industry is an important chemical process to produce chlorine and sodium hydroxide. The most common method involves the electrolysis of a concentrated aqueous solution of sodium chloride (NaCl), which is known as brine.

The chemical reactions that take place in this process are:

 $egin{aligned} 2\mathrm{Cl}^{\ }_{\mathrm{(aq)}} &
ightarrow \mathrm{Cl}_{2(\mathrm{g})} + 2\mathrm{e}^{-} \ 2\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})} + 2\mathrm{e}^{-} &
ightarrow \mathrm{H}_{2(\mathrm{g})} + \ 2\mathrm{OH}^{-}_{\mathrm{(aq)}} \end{aligned}$

There are also Na^+ ions from the NaCl in the solution. If the products are kept separate (to prevent a reaction between the chlorine and hydroxide) the Na^+ will react with the hydroxide ions making the overall reaction as follows: $2NaCl_{(aq)} + 2H_2O_{(l)} \rightarrow Cl_{2(g)} + H_{2(g)} + 2NaOH_{(aq)}$

The uses of chlorine include:

- the purification of water
- as a disinfectant
- in the production of:
 - hypochlorous acid (used to kill bacteria in drinking water), chloroform, carbon tetrachloride
 - paper, food
 - antiseptics, insecticides, medicines, textiles
 - paints, petroleum products, solvents, plastics (such as polyvinyl chloride).

The uses of sodium hydroxide include:

- making soap and other cleaning agents
- purification of bauxite (the ore of aluminium)
- making paper
- making rayon (artificial silk).

If the chlorine and hydroxide ions are not kept separate (i.e. are allowed to react), and the temperature is kept below , then the following occurs:

 $\mathrm{Cl}_{2(g)} + \ 2\mathrm{OH^{\text{-}}}_{(\mathrm{aq})} \rightarrow \mathrm{Cl^{\text{-}}}_{(\mathrm{aq})} + \ \mathrm{ClO^{\text{-}}}_{(\mathrm{aq})} + \mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}$

The balanced chemical equation with the inclusion of the $_{Na^{+}}$ ions is: $\rm Cl_{2(g)}+~2NaOH_{(aq)}\rightarrow NaCl_{(aq)}+~NaClO_{(aq)+}H_2O_{(l)}$

The uses of NaClO (sodium hypochlorite) include:

- use in bleaches, disinfectants, and water treatments
- use during root canal surgery and to neutralise nerve agents.

If the temperature is above and mixing occurs, then: $3Cl_{2(g)}+~6OH^{\text{-}}{}_{(aq)}\rightarrow~5Cl^{\text{-}}{}_{(aq)}+~ClO_{3(aq)}+3H_2O_{(l)}$

And the balanced chemical equation including $_{Na^+}$ ions is: $\rm 3Cl_{2(g)}+~6NaOH_{(aq)}\rightarrow 5NaCl_{(aq)}+~NaClO_{3(aq)}+~3H_2O_{(l)}$

The uses of $NaClO_3$ (sodium chlorate) include:

- making chlorine dioxide
- a herbicide
- generating oxygen in chemical oxygen generators.

If calcium chloride was used instead of sodium chloride, then the products would be calcium chloride, calcium hypochlorite and calcium chlorate. Similarly, if potassium chloride was used, the products would be potassium chloride, potassium hypochlorite and potassium chlorate.

The chemistry of hair and hair products

Hair is primarily made up of a protein called keratin. Keratin is a macromolecule and consists largely of the amino acid cystine (an organic acid). Proteins give hair a natural acidic pH of between 4 and 5. The following hair products alter this pH and do controlled damage to get the desired effect.

- Permanent waving application: Another name for a permanent waving application is a perm. Ammonium thioglycolate is also known as a perm salt. The reversible reaction of ammonium thioglycolate (reactant) to form thioglycolic acid and free ammonia (product) is given below: $HSCH2COO^{-}_{(aq)} + NH^{+}_{4(aq)} \rightleftharpoons HSCH_{2}COOH_{(aq)} + NH_{3(aq)}$
- Hair relaxers: Hair relaxers work by doing controlled damage to your hair in much the same way as perms do. They reduce the cystine to cysteine and allow the hair structure to be changed.

There are two main types of hair relaxers: lye and no-lye. The lye hair relaxers contain sodium hydroxide (NaOH) and have a very high pH (in the range 12 - 14). These types of hair relaxers break bonds and strip the hair of its natural oils. Often heating is also used, and this damages the hair even more.

Due to increasing worry about the damage caused by lye hair relaxers, companies have produced nolye hair relaxers. These still contain basic compounds and are still caustic, but their pH is not as high (pH in the range 9 - 11).

Summary

In this unit you have learnt the following:

- The Arrhenius definition of acids and bases defines an acid as a substance that increases the concentration of hydronium ions (H_3O^+) in a solution. A base is defined as a substance that increases the concentration of hydroxide ions (OH^-) in a solution. However, this definition only applies to aqueous solutions (in water).
- The Brønsted-Lowry definition is much broader. An acid is a substance that donates protons ($_{\rm H^+}$), and a base is a substance that accepts protons.
- In different reactions, certain substances can act as both an acid and a base. These substances are amphoteric substances. Amphiprotic substances are amphoteric substances that are Brønsted-Lowry acids and bases. Water is both amphoteric and amphiprotic.
- A large percentage of molecules in a strong acid or base dissociate or ionise to form ions in solution.
- Only a small percentage of molecules in a weak acid or base dissociate or ionise to form ions in solution.
- In a concentrated solution there is a high ratio of dissolved substance to solvent.
- In a dilute solution there is a low ratio of dissolved substance to solvent.

- When an acid and a base react, they form a salt and water. The salt is made up of a cation from the base and an anion from the acid. An example of a salt is sodium chloride (NaCl), which is the product of the reaction between sodium hydroxide (NaOH) and hydrochloric acid (HCl).
- The reaction between an acid and a base is a neutralisation reaction.
- In the reaction between an acid and a metal the products are a salt and hydrogen.
- In the reaction between an acid and a metal hydroxide or metal oxide the products are a salt and water.
- In the reaction between an acid and a metal carbonate or metal hydrogen carbonate the products are a salt, water, and carbon dioxide.
- The pH scale is a measure of the acidity or alkalinity of a solution. It ranges from 0 to 14. Values greater than 7 indicate a base, while those less than 7 indicate an acid.
- An indicator is a compound that is a different colour in a basic solution, an acidic solution, and at the endpoint of a reaction. They are used to determine the endpoint during a neutralisation reaction.
- Titration is the method used to determine the concentration of a known substance using another, standard, solution. Acid-base titrations are an example.
- Two notable applications of acids and bases are in the chloralkali industry, and in hair products including permanent waving applications, hair relaxers, and hair dyes.

Unit 1: Assessment

Suggested time to complete: 30 minutes

- 1. Calculate the concentration of the following solutions:
 - a. 32.8 mg of hydrochloric acid (HCl) added to 12.76 cm^3 of water
 - b. 1.12 dm^3 of a 6.54 mol.dm^{-3} concentration solution of ammonia (NH₃) added to 0.50 dm^3 of water.
- 2. 25.0 cm^3 of a 0.1 mol.dm^{-3} standard solution of sodium carbonate was used to neutralise 35.0 cm^3 of a solution of hydrochloric acid.
 - a. Write a balanced chemical equation for the reaction.
 - b. Calculate the concentration of the acid.
- 3. The stomach secretes gastric juice, which contains hydrochloric acid. The gastric juice helps with digestion. Sometimes there is an overproduction of acid, leading to heartburn or indigestion. Antacids, such as milk of magnesia, can be taken to neutralise the excess acid. Milk of magnesia is only slightly soluble in water and has the chemical formula $Mg(OH)_2$.
 - a. Write a balanced chemical equation to show how the antacid reacts with the acid.
 - b. The instructions on the bottle recommend that children under the age of 12 years take one teaspoon of milk of magnesia, whereas adults can take two teaspoons of the antacid. Briefly explain why the dosages are different.
 - c. Why is it not advisable to take an overdose of the antacid? Refer to the hydrochloric acid concentration in the stomach in your answer.
- 4. A certain antacid tablet contains 22.0 g of baking soda (NaHCO₃). It is used to neutralise the excess hydrochloric acid in the stomach. The balanced equation for the reaction is: $NaHCO_{3(s)} + HCl_{(aq)} \rightarrow NaCl_{(aq)} + H_2O_{(l)} + CO_{2(g)}$ The hydrochloric acid in the stomach has a concentration of 1.0 mol.dm⁻³.

Calculate the volume of the hydrochloric acid that can be neutralised by the antacid tablet.

- 5. A learner finds some sulfuric acid solution in a bottle labelled 'dilute sulfuric acid'. He wants to determine the concentration of the sulfuric acid solution. To do this, he decides to titrate the sulfuric acid against a standard potassium hydroxide (KOH) solution.
 - a. What is a standard solution?

- b. Calculate the mass of KOH which he must use to make $_{300}$ cm 3 of a $_{0.2}$ mol.dm $^{-3}$ KOH solution.
- c. Write a balanced chemical equation for the reaction between $\mathrm{H}_2\mathrm{SO}_4$ and $\mathrm{KOH}.$
- d. During the titration the learner finds that $_{15}$ cm³ of the solution KOH neutralises $_{20}$ cm³ of the H_2SO_4 solution. Calculate the concentration of the H_2SO_4 solution.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

٦.

- a. Concentrated. There are many more moles of the base than the solvent.
- b. Dilute. There are many more moles of the solvent than the acid.

2.

a. You have the volume (V) and the mass (m). You need the number of moles (n) and the molar mass (M).

$$egin{aligned} M_{\mathrm{(NaHCO_3)}} &= 23 + 1.01 + 12 + (3 ext{ x } 16) = 84.01 ext{ g.mol}^{-1} \ n_{\mathrm{(NaHCO_3)}} &= rac{m}{M} = rac{27}{84.01} = 0.32 ext{ mol} \ V &= 22, 6 ext{ cm}^3 imes rac{0.001}{1} = 0.0226 ext{ dm}^3 \ c_{\mathrm{(NaHCO_3)}} &= rac{n}{V} = rac{0.32}{0.0226} = 14.16 ext{ mol.dm}^{-3} \end{aligned}$$

b. You have the volume (V) and number of moles (n). You can calculate the concentration.

$$c_{(H_3PO_4)} = rac{n}{V} = rac{0.893}{4.79} = 0.186 \,\, {
m mol.dm^{-3}}$$

Back to Exercise 1.1

Exercise 1.2

- 1. Acid (HCl) + metal (Ca) \rightarrow salt + hydrogen Anion (from acid) is Cl⁻, cation (from metal) isCa²⁺. Therefore salt isCaCl₂. HCl_(aq) + Ca_(s) \rightarrow CaCl_{2(aq)} + H_{2(g)} To balance this equation there must be two HCl molecules. 2HCl_(aq) + Ca_(s) \rightarrow CaCl_{2(aq)} + H_{2(g)}
- $\begin{array}{ll} \text{2. acid (HCl) + metal hydroxide } (Mg(OH)_2) \rightarrow \text{salt + water} \\ \text{Anion (from acid) is Cl^-, cation (from metal) is } Mg^{2+}. \\ \text{Therefore salt is} MgCl_2. \\ \text{HCl}_{(aq)} + Mg(OH)_{2(s)} \rightarrow MgCl_{2(aq)} + H_2O_{(l)} \\ \text{To balance this equation there must be two } HCl \\ \text{molecules and two } H_2O \\ \text{molecules.} \\ 2HCl_{(aq)} + Mg(OH)_{2(s)} \rightarrow MgCl_{2(aq)} + 2H_2O_{(l)} \\ \end{array}$
- 3. acid (H₂SO₄) + metal oxide (MgO) → salt + water Anion (from acid) is SO₂⁴⁻, cation (from metal) is Mg²⁺. Therefore salt is MgSO₄. H₂SO_{4(aq)}+MgO_(s) → MgSO_{4(aq)}+ H₂O_(l) This equation is balanced.

Exercise 1.3

1. We need the mass of oxalic acid. However, we don't know the number of moles yet.

 $egin{aligned} V &= 500 \,\, {
m cm}^3 imes rac{0.001}{1} = 0.5 \,\, {
m dm}^3 \ n &= c {
m x} V \ n &= 0.2 \,\, {
m x} \,\, 0.5 = 0.1 \,\, {
m mol} \ M_{
m (H_2C_2O_4)} &= (2 {
m x} 1.01 + 2 {
m x} 12, 0 + 4 {
m x} 16.0) = 90.02 \,\, {
m g.mol}^{-1} \ m &= n M \ m &= 0.1 \,\, {
m x} \,\, 90.02 = 9.00 \,\, {
m g} \end{aligned}$

2. The balanced equations is:

 $\begin{array}{l} \mathrm{H}_{2}\mathrm{C}_{2}\mathrm{O}_{4}\ +\ 2\mathrm{NaOH} \rightarrow \mathrm{Na}_{2}\mathrm{C}_{2}\mathrm{O}_{4}\ +\ 2\mathrm{H}_{2}\mathrm{O}\\ \text{The number of moles of oxalic acid used is the number of moles in }_{40\ \mathrm{cm}^{3}} \text{ of the standard solution:}\\ V = 40\ \mathrm{cm}^{3}\ \times\ \frac{0.001}{1} = 0.04\ \mathrm{dm}^{3}\\ n = \mathrm{cx}V = 0.2\ \mathrm{x}\ 0.04 = 0.008\ \mathrm{mol} \end{array}$

The molar ratio of oxalic acid to sodium hydroxide is
$$1 : 2$$
. For everyone mole of oxalic acid there are two moles of sodium hydroxide.

$$egin{aligned} n_{\mathrm{(NaOH)}} &= 2 imes 0,008 = 0.016 \,\,\mathrm{mol} \ V_{\mathrm{(NaOH)}} &= 35\,\,\mathrm{cm}^3 imes rac{0,001}{1} = 0.035\,\,\mathrm{dm}^3 \ c_{\mathrm{(NaOH)}} &= rac{n}{V} = rac{0.016}{0.035} = 0.46\,\,\mathrm{mol.dm}^{-3} \end{aligned}$$

Back to Exercise 1.3

Unit 1: Assessment

1.

a. You have the volume (V) and the mass (m). You need the number of moles (n) and the molar mass (M).

$$egin{aligned} M_{
m (HCl)} &= 1.01 + 35.45 = 36.46 \,\, {
m g.mol}^{-1} \ m_{
m (HCl)} &= 32.8 \,\, {
m mg} \,\, imes rac{0.001}{1} \,\, = 0.0328 \,\, {
m g} \ n_{
m (HCl)} &= rac{m}{M} = rac{0.0328}{36.46} = 9.0 imes 10^{-4} \,\, {
m mol} \ V &= 12,76 \,\, {
m cm}^3 \,\, imes rac{0.001}{1} = 0.01276 \,\, {
m dm}^3 \ c_{
m (HCl)} &= rac{n}{V} = rac{9.0 imes 10^{-4}}{0.01276} = \,\, 0.07 \,\, {
m mol.dm}^{-3} \end{aligned}$$

b. You have the volume (V) and original concentration of the ammonia solution, you need the number of moles (n) in that original solution.

$$c = \frac{n}{V}$$
. Therefore $n = c \ge V$

 $n_{
m (original \ solution \ of \ NH_3)} = 6.54 \ {
m x} \ 1,\!12 = 7.32 \ {
m mol}$

The final volume used to calculate the concentration must be the volume of the water plus the volume of ammonia added.

$$V_{
m (total)} = 0.50 + 1.12 = 1.62 \,\, {
m dm}^3$$
 $c_{
m (final solution of NH_3)} = rac{n}{V} = rac{7.32}{1.62} = 4.52 \,\, {
m mol.dm}^{-3}$

2.

a. metal carbonate + acid + salt + carbon dioxide + water The cation will be from the metal (Na^+) , the anion will be from the acid (CI^-) . Therefore the salt will be NaCl. Ν

$$\mathrm{Na_2CO}_{3(\mathrm{aq})} + 2\mathrm{HCl}_{(\mathrm{aq})}
ightarrow 2\mathrm{NaCl}_{(\mathrm{aq})} + \mathrm{CO}_{2(\mathrm{g})} + \mathrm{H_2O}_{(l)}$$

b.

 $n = \mathrm{c} imes V$ $V_{
m (Na_2CO_3)} = 25.0~{
m cm}^3 imes {0.001 \over 1} = 0.0250~{
m dm}^3$ $n_{({
m Na}_2{
m CO}_3)}=c\,\,{
m x}\,\,V=0.1{
m x}\,\,0.0250=0.0025\,\,{
m mol}$ From the balanced equation for every $1 ext{ mol of Na}_2 ext{CO}_3$ there are $2 ext{ mol HCl}$. $n_{
m (HCl)} = 2 \, \, {
m x} \, \, 0.0025 = 0.005 \, \, {
m mol}$ $V_{
m (HCl)} = 35.0~{
m dm}^3 imes {0.001\over 1} = 0.0350~{
m dm}^3$ $c_{
m (HCl)} = rac{n}{V} = rac{0.005}{0.0350} = 0.143 \; {
m mol.dm^{-3}}$

3.

- a. base + acid → salt + water The cation (from the base) is Mg^{2+} . The anion (from the acid) is CI^- . Therefore the salt is $MgCl_2$ $\mathrm{Mg(OH)}_{2(\mathrm{s})}\!+\!2\mathrm{HCl}_{(\mathrm{aq})}
 ightarrow\mathrm{MgCl}_{2(\mathrm{aq})}\!+\!2\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}$
- b. Adults have a bigger mass and generally produce more acid than children. Adults will therefore need more antacid to neutralise the excess acid. If children were to take the same dosage as adults, they would have excess base in their stomach.
- c. A low acid concentration (pH too high) in the stomach may slow down food digestion or may cause further stomach upset.

4.

 $M_{
m (NaHCO_3)} = 23.0\,+\,1.01\,+\,12.0\,+\,(3\,\,{
m x}\,\,16.0)\,=\,84.01\,\,{
m g.mol}^{-1}$

$$n_{
m (NaHCO_3)} = rac{m}{M} \ n = rac{22.0}{84.01} = 0.26 \,\, {
m mol}$$

From the balanced equation we see that the molar ratio of NaHCO₃ to HCl is 1:1. Therefore:

$$egin{aligned} n_{(ext{HCl})} &= n_{(ext{NaHCO}_3)} = 0.26 \ ext{mol} \ c &= rac{n}{V} \ ext{therefore} \ V = rac{n}{c} \ V_{(ext{HCl})} &= rac{0.26}{1.0} \ = 0.26 \ ext{dm}^3 \end{aligned}$$

5.

a. A standard solution is a solution that contains a precisely known concentration of a substance. This substance can then be used in titrations.

b.

 $V_{
m (KOH)} = 300 \,\,{
m cm}^3 \,\, imes rac{0.001}{1} = 0,3 \,\,{
m dm}^3$ $n = c \ge V$ $n_{(\text{KOH})} = 0.2 \ge 0.3 = 0.06 \text{ mol}$ $M_{
m (KOH)} = (39.1 \ + \ 16.0 \ + \ 1.01) = 56.11 \ {
m g.mol}^{-1}$ $m = n \mathbf{x} M$ $m_{(\text{KOH})} = 0.06 \text{ x } 56.11 = 3.37 \text{ g}$

c. Sulfuric acid is a strong acid and potassium hydroxide is a strong base, therefore the equation is: $\mathrm{H}_2\mathrm{SO}_{4(\mathrm{aq})}\!+\!2\mathrm{KOH}_{(\mathrm{aq})}
ightarrow\mathrm{K}_2\mathrm{SO}_{4(\mathrm{aq})}\!+\!2\mathrm{H}_2\mathrm{O}_{(\mathrm{l})}$

d.

$$V_{
m (KOH)} = 15\,\,{
m cm}^3~ imes {0,001\over 1} = 0.015\,\,{
m dm}^3$$

The number of moles of KOH used to neutralise the $\rm H_2SO_4$ is: $n_{\rm (KOH)} = \rm c~x~V = 0.2~x~0.015 = 0.003~mol$

From the balanced equation we see that the mole ratio of H_2SO_4 to KOH is 1:2. There is one mole of sulfuric acid for every two moles of potassium hydroxide.

$$\begin{split} n_{\rm (H_2SO_4)} &= \frac{0.003}{2} = 0.0015 \,\, {\rm mol} \\ V_{\rm (H_2SO_4)} &= 20 \,\, {\rm cm}^3 \,\, \times \frac{0.001}{1} \, 0.020 \,\, {\rm dm}^3 \\ \text{Therefore, the concentration of the sulfuric acid is:} \\ c_{\rm (H_2SO_4)} &= \frac{n}{V} = \frac{0.0015}{0.020} = 0,075 \,\, {\rm mol.dm}^{-3} \end{split}$$

Back to Unit 1: Assessment

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SUBJECT OUTCOME XII CHEMICAL CHANGE: IDENTIFY AND APPLY KNOWLEDGE OF ELECTROCHEMICAL CELLS

Subject outcome

Subject outcome 6.2: Identify and apply knowledge of electrochemical cells



*

Learning outcomes

- Define oxidation, reduction, cathode, anode, and electron transfer.
- Write oxidation and reduction half reactions.
- Describe the purpose of a sacrificial anode and corrosion of metals and write reactions down to illustrate.
- Describe an electrochemical cell and identify examples in everyday life/industry.
- Draw and label an electrochemical cell and determine cell potential and compare with cells used in everyday life (Li, Cd, Pb- cells).



Unit 1 outcomes

By the end of this unit you will be able to:

- Define oxidation, reduction, cathode, anode, and electron transfer.
- Write oxidation and reduction half reactions.
- Describe the purpose of a sacrificial anode and corrosion of metals and write reactions down to illustrate.
- Describe an electrochemical cell and identify examples in everyday life/industry.
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Unit 1: Electrochemical Cells

EMMA HARRAGE



Unit outcomes

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- Describe an electrochemical cell and identify examples in everyday life/industry.
- Draw and label an electrochemical cell and determine cell potential and compare with cells used in everyday life (Li, Cd, Pb- cells).

What you should know

Before you start this unit, make sure you:

- Can balance chemical equations. Refer to level 3 subject outcome 6.2 unit 1 if you need help with this.
- Understand oxidation numbers. Refer to level 3 subject outcome 5.2 unit 3 if you need help with this.
- Understand that there are different types of chemical reactions. Refer to <u>level 3 subject outcome 6.3</u> <u>unit 1</u> if you need help with this.

Introduction

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In this unit you will learn about electrolysis which is what happens inside an electrolytic cell. Electrolysis is a process where electricity is used to make a chemical change happen. The most common usage is for decomposition. Sodium chloride is broken down by electrolysis to obtain sodium and chlorine. Sodium chloride must be melted until it is liquid and then the Na^+ ions and the Cl⁻ can move to the oppositely charge anode.

There are two types of electrochemical cell – electrolytic in which electricity is used for decomposition, and galvanic which are used to make electricity, both involving redox reactions. Electrochemical cells use metallic electrodes in an electrolyte solution to generate a current and are called galvanic cells. Batteries usually have at least two of these types of cells.

Redox reactions

Redox reactions involve the exchange of electrons. One species loses electrons and becomes more positive, while the other species gains electrons and becomes more negative. To decide if a redox reaction has occurred, we look at the charge of the atoms, ions or molecules involved. If one of them has become more

positive and the other one has become more negative, then a redox reaction has occurred. For example, sodium metal is oxidised to form sodium oxide and sometimes sodium peroxide as well. The balanced equation for this is:

 $4\mathrm{Na}{+}\mathrm{O_2} \rightarrow 2\mathrm{Na_2O}$

In the above reaction sodium and oxygen are both neutral and so have no charge. In the products however, the sodium atom has a charge of 1^+ and the oxygen atom has a charge of 2^- . This tells us that the sodium has lost electrons and the oxygen has gained electrons. Since one species has become more positive and one more negative, we can conclude that a redox reaction has occurred. We could also say that electrons have been transferred from one species to the other. Redox is short for oxidation and reduction.





Figure 1: This mnemonic, OIL RIG, is an easy way to remember redox



Figure 2: In both oxidation and reduction, a transfer of electrons is involved resulting in a change in the oxidation state of the elements

An element or compound that loses electrons is oxidised.

e.g. $\operatorname{Zn}_{(s)}
ightarrow \operatorname{Zn}^{2+}{}_{(aq)} + 2e^{-2\pi i t}$

As it loses electrons it gives them away to another element or compound and the element or compound it gives the electrons to is reduced. This makes the compound or element which loses electrons a reducing agent.

An element or compound that gains electrons is reduced.

e.g.
$$\mathrm{Cu}^{2+}{}_{\mathrm{(aq)}}$$
 + 2e \rightarrow $\mathrm{Cu}_{\mathrm{(s)}}$

As it gains electrons it takes them away from another element or compound and the element or compound it takes them from is oxidised. This makes the compound or element which gains electrons an oxidising agent.



- a. oxidation
- b. reduction
- c. oxidising agent
- d. reducing agent
- 2. In each of the following reactions say whether the reactant iron species (Fe, Fe^{2+} , Fe^{3+}) is oxidised or reduced.
 - a. $\mathrm{Fe}_{(s)} \rightarrow \mathrm{Fe}^{2+}{}_{(\mathrm{aq})}{+}2\mathrm{e}^{-}$
 - b. $Fe_2O_{3(s)} + 2Al_{(s)} \rightarrow Al_2O_{3(s)} + 2Fe_{(s)}$

The full solutions are at the end of the unit.

Writing redox reactions using half-reactions

Remember that oxidation and reduction occur simultaneously in a redox reaction. The reactions taking place in electrochemical cells are redox reactions. Two questions should be asked to determine if a reaction is a redox reaction:

- Is there an atom or ion being oxidised?
- Is there an atom or ion being reduced?

If the answer to both of these questions is yes, then the reaction is a redox reaction. For example, this reaction is a redox reaction:

 $2 Fe^{3+}_{(aq)} + Sn^{2+}_{(aq)}
ightarrow \ 2 Fe^{2+}_{(aq)} + Sn^{4+}_{(aq)}$

You can write a redox reaction as two half-reactions, one showing the reduction process, and one showing the oxidation process. Fe^{3+} is gaining an electron to become Fe^{2+} . Iron (III) ions are therefore being reduced and Sn^{2+} tin is the reducing agent (causing iron to be reduced).

The reduction half-reaction is: $Fe^{3+}_{(aq)}\,+\,e^{\text{-}}\rightarrow Fe^{2+}_{(aq)}$

 s_n^{2+} is losing two electrons to become s_n^{4+} . Tin (II) ions are therefore being oxidised and iron (III) ions are the oxidising agent (causing s_n^{4+} to be oxidised).

The oxidation half-reaction is: $Sn^{2+}_{(aq)} \rightarrow Sn^{4+}_{(aq)} \, + \, 2e^{-}$

Notice that in the overall reaction, the reduction half-reaction is multiplied by two. This is so that the number of electrons gained in the reduction half-reaction match the number of electrons lost in the oxidation half-reaction. Half-reactions can be used to balance redox reactions.



Step 1: Write down the unbalanced oxidation half-reaction

$$\mathrm{Fe}^{2+}_{\mathrm{(aq)}}
ightarrow \mathrm{Fe}^{3+}_{\mathrm{(aq)}}$$

Step 2: Balance the number of atoms on both sides of the equation

There is one iron atom on the left and one on the right, so no additional atoms need to be added. Once the atoms are balanced, check that the charges balance.

The charge on the left of the equation is +2, but the charge on the right is +3. Therefore, one electron must be added to the right-hand side so that the charges balance.

The half-reaction is now: ${\rm Fe}^{2+}_{(aq)}
ightarrow {\rm Fe}^{3+}_{(aq)} + {\rm e}^-$

Step 3: Repeat steps 1 to 3 with the reduction half-reaction

The unbalanced reduction half-reaction is: ${\rm Cl}_{2(g)} \to {\rm Cl}_{({\rm aq})}$

The atoms don't balance, so we need to multiply the right-hand side by two to fix this. $\rm Cl_{2(g)}\to 2\rm Cl_{(aq)}$

Step 4: Two electrons must be added to the left-hand side to balance the charges

$$\mathrm{Cl}_{2(g)}{+}2e^{\text{-}} \rightarrow 2\mathrm{Cl}_{(aq)}^{\text{-}}$$

Step 5: Compare the number of electrons in each equation

Multiply each half-reaction by a suitable number so that the number of electrons released (oxidation) is equal to the number of electrons accepted (reduction).

Oxidation half-reaction: x 2: $2Fe^{2+}_{(aq)} \rightarrow 2Fe^{3+}_{(aq)} + 2e^{-}$

Reduction half-reaction: x 1: $Cl_{2(g)} + 2e^- \rightarrow 2Cl_{(aq)}^-$

Step 6: Combine the two half-reactions to get a final equation for the overall reaction

Note that the electrons on each side of the redox equation will cancel each other out.

$$2Fe^{2+}_{(aq)}+\ Cl_{2(g)}\rightarrow 2Fe^{3+}_{(aq)}+\ 2Cl^{-}_{(aq)}$$

Step 7: Do a final check to make sure that the equation is balanced

We check the number of atoms and the charges and find that the equation is balanced.



Exercise 1.2

1. Balance the following redox equations in an acidic solution:

a.
$$\mathrm{H^+} + \mathrm{Mn^{2+}} + \mathrm{BiO_3^-} \rightarrow \mathrm{MnO_4^-} + \mathrm{Bi^{3+}} + \mathrm{H_2O}$$

- b. $\mathrm{H_2O}\,+\,\mathrm{P}\,+\,\mathrm{Cu}^{2+}\rightarrow\mathrm{Cu}\,+\,\mathrm{H_2PO_4^{\text{-}}}\,+\,\mathrm{H^+}$
- $\text{c.} \hspace{0.1in} H_2O \hspace{0.1in} + \hspace{0.1in} NO_2 \rightarrow NO_3^{-} \hspace{-0.1in} + \hspace{-0.1in} NO \hspace{0.1in} + \hspace{-0.1in} H^+$

2. Balance the following redox equation in a basic solution:

a.
$$Al^{3+} + H_2O + OH^- \rightarrow AlO_2 + H_2$$

b.
$$OH^- + Pb^{2+} + IO_3^- \rightarrow H_2O + PbO_2 + I_2$$

The <u>full solutions</u> are at the end of the unit.

Electrolysis

lonic substances contain charged particles called ions. For example, lead bromide contains positively charged lead ions and negatively charged bromide ions. Electrolysis is the process by which ionic substances are decomposed (broken down) into simpler substances when an electric current is passed through them. For electrolysis to work, the compound must contain ions. Covalent compounds cannot act as electrolytes because they contain neutral atoms.

To be an electrolyte, a substance must be able to conduct electricity. To test for conductivity, include a sample of the substance in a circuit that contains a source of direct current and a bulb or an ammeter. If a current flows through the substance when it is molten or dissolved, it is an electrolyte.

The ions must be free to move, which is possible when an ionic substance is dissolved in water, or it is melted. For example, if electricity is passed through molten lead bromide, the lead bromide is broken down to form lead and bromine.

This is what happens during electrolysis:

- Positively charged ions move to the negative electrode, the cathode, during electrolysis. They receive electrons and are reduced.
- Negatively charged ions move to the positive electrode, the anode, during electrolysis. They lose electrons and are oxidised. The substance that is broken down is called the electrolyte.



Figure 3: Electrolysis of copper chloride

Two copper electrodes are placed in a solution of blue copper chloride and are connected to a source of electrical current. The current is turned on for a period of time. At the positive anode, copper metal is oxidised to form Cu^{2+} ions. This is why it appears that the copper has dissolved from the electrode. The reaction at this electrode is:

 $\mathrm{Cu}_{(\mathrm{s})}
ightarrow \mathrm{Cu}^{2+}{}_{(\mathrm{aq})} + 2\mathrm{e}^{-}$

Electrolysis of molten salts

New substances form when a molten or dissolved ionic compound conducts electricity. For example, lead bromide can be used to produce bromine and lead using the following process:



Figure 4: The electrolysis of lead bromide

Solid lead (II) bromide is placed in a crucible and heated over a Bunsen burner until it melts. Two carbon electrodes are inserted into the molten electrolyte and a direct current is passed between them. Carbon electrodes are chosen because they have a high melting point and are inert.

The observations are summarised in the table:

Observation	Explanation
Anode (positive)	Bubbles of brown gas Bromine vapour (Br_2) is produced. $2Br_{(aq)} \rightarrow \ Br_2^{2+}{}_{(aq)} + 2e^-$
Cathode (negative)	Silvery liquid metal present at this electrode underneath the remaining molten electrolyte. Lead (Pb) is produced. $Pb_{(aq)}^{2^+} + 2e^- \rightarrow Pb_{(aq)}$

It is easy to predict the products of electrolysis of molten electrolytes because they simply split into their elements:

- the metal is formed at the cathode (negative electrode) because that is where the positive metal ions are attracted.
- the non-metal element is formed at the anode (positive electrode) where the negative non-metal ions are attracted.

Compound	Anode	Cathode
Zinc Chloride	Chlorine	Zinc
Aluminium Oxide	Oxygen	Aluminium

Note

Spectator ions may be either cations (positively charged ions) or anions (negatively charged ions). The ion is unchanged on both sides of a chemical equation and does not affect equilibrium. When writing a net ionic equation, spectator ions found in the original equation are ignored.
Chloralkali industry

Sodium chloride (NaCl), commonly known as salt, is used to treat icy roads and to season food, as well as to produce chemicals for a range of uses. Hydrogen is one of the three products, as well as sodium and chlorine from the electrolysis of sodium chloride solution.

Sodium chloride can be made in a laboratory by reacting sodium with chlorine. However, it is found naturally in large amounts in seawater or in underground deposits. It is often obtained either by evaporating seawater or by mining underground deposits.

Salt is mined by solution mining. Water is pumped down into the salt deposit. Salt dissolves in the water, forming a concentrated salt solution. This is then pumped up to the surface ready for use in the chemical industry. Solution mining is a continuous process that is safer than sending miners underground.

Salt is widely used in the food industry as a preservative and flavour enhancer. It is vital for human health – we need sodium in our diet to allow our bodies to carry out essential functions. Sodium chloride is the raw material for the manufacture of hydrogen, chlorine, and sodium hydroxide using electrolysis. The electrolysis of sodium chloride solution (brine) is an important industrial process because hydrogen, chlorine and sodium hydroxide have many uses.



Figure 5: When an electric current is passed through concentrated sodium chloride solution, hydrogen gas forms at the negative electrode, chlorine gas forms at the positive electrode, and a solution of sodium hydroxide also forms

You might have expected sodium metal to be deposited at the negative electrode. However, sodium ions are spectator ions and do not react. The water in the solution reacts to produce hydrogen. and hydroxide ions.

Since chloride ions are removed and hydroxide ions are produced by the electrolysis, the electrolyte gradually changes from a solution of sodium chloride to a solution of sodium hydroxide.

During electrolysis:

Hydrogen ions $H^+_{(aq)}$ (from the water) are discharged at the cathode as hydrogen gas, $H_{2(g)}$: $2H_2O_{(l)}+2e^- \rightarrow H_{2(g)}+2OH^-_{(aq)}$

 $\text{Chloride ions } \operatorname{Cl}_{(\mathrm{aq})} \text{ are discharged at the anode as chlorine gas, } \operatorname{Cl}_{2(g)}: 2\operatorname{Cl}_{_{(\mathrm{aq})}} \to \operatorname{Cl}_{2(g)} + 2e$

Sodium ions $Na^+_{(aq)}$ and hydroxide ions $OH^-_{(aq)}$ (from the water) stay behind – they form sodium hydroxide solution, $NaOH_{(aq)}$

The overall reaction is: $2H_2O_{(l)}+2Cl_{(aq)}\rightarrow H_{2(g)}+Cl_{2(g)}+2OH_{(aq)}$

The three products of the electrolysis of concentrated sodium chloride solution have important uses in the chemical industry:

- hydrogen is used as a fuel and for making ammonia
- · chlorine is used to sterilise water supplies, and to make bleach and hydrochloric acid
- sodium hydroxide is used to make soap, paper, and bleach.

Electrolytic cells

In an electrolytic cell electrical potential energy is converted to chemical potential energy. An electrolytic cell uses an electric current to force a particular chemical reaction to occur, which would otherwise not take place.

Sometimes galvanic cells are called electrochemical cells. While they are electrochemical cells, electrolytic cells are also electrochemical cells. Electrolytic and galvanic cells are not the same, however.

An electrolytic cell is activated by applying an electrical potential across the electrodes to force an internal chemical reaction between the electrodes and the ions that are in the electrolyte solution. This process is called electrolysis.



Figure 6: A sketch of an electrolytic cell

In an electrolytic cell:

- the electrolyte solution consists of the metal cations and spectator anions
- the oxidation and reduction reactions occur in the same container but are non-spontaneous; they require the electrodes to be connected to an external power source to proceed
- the electrodes in an electrolytic cell can be the same metal or different metals; the principle is the same: let there be only one metal, and let that metal be Z
- an electrode is connected to the positive terminal of the battery; oxidation is loss at the anode
- to balance the charge at the anode metal atoms are oxidised to form metal ions; the ions move into solution, leaving their electrons on the electrode
- + the following reaction takes place: $Z_{(s)} \rightarrow Z^+_{(aq)} + e^-$
- an electrode is connected to the negative terminal of the battery; reduction is gain at the cathode
- when positive ions come in contact with the cathode, the ions gain electrons and are reduced

- + the following reaction takes place: $Z^+_{(aq)} + e^{\text{-}} \to Z_{(s)}$
- + this means that the overall reaction is: $Z_{(s)} + \, Z^+_{(aq)} \to Z^+_{(aq)} + \, Z_{(s)}.$

While this might seem trivial, this is an important technique to purify metals. Water can undergo electrolysis to form hydrogen gas and oxygen gas according to the following reaction: $2H_2O_{(l)} \rightarrow 2H_{2(g)}+O_{2(g)}$.



Figure 7: The decomposition of water by electrolysis

This reaction is very important because hydrogen gas has the potential to be used as an energy source. The electrolytic cell for this reaction consists of two electrodes, submerged in an electrolyte, and connected to a source of electric current.

The oxidation half-reaction is as follows: $4OH^{-}{}_{(aq)}
ightarrow 2H_2O_{(l)} + O_{2(g)} + 4e^{-}$

The reduction half-reaction is as follows: $4 H^+{}_{(aq)} + 4 e^-
ightarrow 2 H_{2(g)}$

Galvanic cells

A galvanic cell (which is also sometimes referred to as a voltaic or wet cell) consists of two half-cells, which convert chemical potential energy into electrical potential energy. In a galvanic cell there are two half-cells. Each half-cell contains an electrode in an electrolyte. The separation is necessary to prevent direct chemical contact of the oxidation and reduction reactions, creating a potential difference. The electrons released in the oxidation reaction travel through an external circuit (and do work) before being used by the reduction reaction.



Figure 8: A sketch of a galvanic cell



Figure 9: A sketch of the electrodes in a galvanic cell

The metal at the anode is X. Oxidation is loss of electrons at the anode.

The anode half-reaction is: $X_{(s)} \rightarrow X^+_{(aq)} \, + \, e^{\text{-}}$

This half-reaction occurs in the half-cell containing the $X_{(s)}$ anode and the $X^+_{(aq)}$ electrolyte solution. The electrons released in the oxidation of the metal remain on the anode, while the metal cations formed move into solution.

The metal at the cathode is Y. Reduction is gain of electrons at the cathode.

The cathode half-reaction is: $Y^+_{(aq)}\,+\,e^{\scriptscriptstyle -}\to Y_{(s)}$

This half-reaction occurs in the half-cell containing the $Y_{(s)}$ cathode and the $Y^+_{(aq)}$ electrolyte solution. At the cathode metal ions in the solution are being reduced (accepting electrons) and deposited on the electrode.

Electrons will flow from areas of high concentration to areas of low concentration, therefore the electrons move from the anode, through the external circuit, to the cathode. Conventional current is measured as a flow of positive charge and so is in the opposite direction (from the cathode to the anode). The overall reaction is: $X_{ij} + Y^+ \rightarrow X^+ + Y_{ij}$

The overall reaction is: $X_{(s)}\!+\!Y^+_{(aq)}\to X^+_{(aq)}\!+\!Y_{(s)}$

To represent this reaction using standard cell notation we write the following: $X_{(s)} |X_{(aq)}^+| |Y_{(aq)}^+| Y_{(s)}$

By convention:

- The anode is always written on the left.
- The cathode is always written on the right.
- The anode and cathode half-cells are divided by || representing the salt bridge.
- The different phases within each half-cell (solid (s) and aqueous (aq) here) are separated by |.
- The electrodes in each half-cell are connected through a wire in the external circuit. There is also a salt bridge between the individual half-cells.

A galvanic cell uses the reactions that take place at the two electrodes to produce electrical energy, i.e. the reaction occurs without the need to add energy.

Note

A galvanic cell generates voltage across their terminals and are a source of energy whilst electrolytic cells need an external potential difference.

	Galvanic cell	Electrolytic cell
Chemical reactions	spontaneous reactions	non-spontaneous reactions
Energy changes	Chemical potential energy from chemical reactions is converted to electrical energy	An external supply of electrical energy causes a chemical reaction to occur
Anode	is negative , oxidation occurs at anode	is positive , oxidation occurs at anode
Cathode	is positive , reduction occurs at cathode	is negative , reduction occurs at cathode
Cell set-up	two half-cells, one electrode in each, connected by a salt-bridge	one cell, both electrodes in cell, no salt-bridge
Electrolyte solution(s)	the electrolyte solutions are kept separate from one another, and are connected by a salt bridge	The cathode and anode are in the same electrolyte
Applications	batteries	electrolysis e.g. of water, NaCl, electroplating

Table 1: A comparison of galvanic and electrolytic cells



Exercise 1.3

- 1. A galvanic cell consists of two half-cells: a copper anode in a copper nitrate $(Cu(NO_3)_{2(aq)})$ solution, and a silver cathode in a silver nitrate $(AgNO_{3(aq)})$ solution.
 - a. Give equations for the half-reactions that take place at the anode and cathode.
 - b. Write the overall reaction for this cell.
 - c. Give standard cell notation for this cell.
 - d. Draw a simple diagram of the galvanic cell. On your diagram, show the direction in which current flows.
- 2. An electrolytic cell consists of two electrodes in a silver chloride (AgCl) solution, connected to a source of current. A current is passed through the solution and Ag^+ ions are reduced to a silver metal deposit on one of the electrodes.
 - a. What is the name of this process?
 - b. Does reduction occur at the electrode where the deposit formed?
 - c. Give the equation for the reduction half-reaction and for the oxidation half-reaction.

The <u>full solutions</u> are at the end of the unit.

Electrochemical reactions

We use batteries throughout our day-to-day lives. Cell phones use lithium-ion batteries, cars use lead-acid batteries, while silver-oxide batteries are used in watches. Some batteries are rechargeable, while others cannot be recharged and have to be thrown away.



Figure 10: Rechargeable batteries

A battery consists of multiple electrochemical cells. And within each cell there are electrochemical reactions taking place. The chemical potential energy can be converted to electrical potential energy, which forms an electric current. The transfer of electrons in a chemical reaction can cause electrical current to flow. If you supply an electric current, it can cause a chemical reaction to take place, by supplying the electrons (and potential energy) necessary for the reactions taking place within the cell.

These types of reactions are called electrochemical reactions.

An electrochemical reaction is a reaction where either:

• a spontaneous chemical reaction creates an electrical potential difference, and therefore an electric current in the external conducting wires

or

• an electric current provides electrical potential energy and electrons, and therefore forces a nonspontaneous chemical reaction to take place. Electrochemistry is the branch of chemistry that studies these electrochemical reactions. An electrochemical cell is a device in which electrochemical reactions take place.

Electrochemical cells

There are two types of electrodes in an electrochemical cell, the anode, and the cathode. The electrode is an electrical conductor that connects the electrochemical species from its solution to the external electrical circuit of the cell.

Oxidation always occurs at the anode while reduction always occurs at the cathode. So when trying to determine which electrode you are looking at, first determine whether oxidation or reduction is occurring there. An easy way to remember this is:



Figure 11: Oxidation is loss of electrons at the anode: an xx; reduction is gain of electrons at the cathode: red cat

The electrode is placed in an electrolyte solution within the cell. An electrolyte is a solution that contains free ions, and which therefore behaves as a conductor of charges (electrical conductor) in solution.

If the cell is made up of two compartments, as in a galvanic cell, those compartments will be connected by a salt bridge. A salt bridge is a material which contains electrolytic solution and acts as a connection between two half-cells (and completes the circuit). It maintains electrical neutrality in and between the electrolytes in the half-cell compartments.

The salt bridge acts as a transfer medium that allows ions to flow through without allowing the different solutions to mix and react directly. It allows a balancing of the charges in the electrolyte solutions and allows the reactions in the cell to continue. Without the salt bridge, the flow of electrons in the outer circuit stops completely. This is because the salt bridge is needed to complete the circuit.

Half-cells

Galvanic cells are made up of two half-cells. A half-cell is a structure that consists of a conductive electrode surrounded by a conductive electrolyte. One half-cell contains the anode and an electrolyte containing the same metal cations. The other half-cell contains the cathode and an electrolyte containing the same metal cations. These half-cells are connected by a salt-bridge and the electrodes are connected through an external circuit.



Figure 12: A zinc and copper galvanic cell

In the diagram of a zinc-copper cell (see figure 12) the cell is made up of a zinc half-cell, containing a zinc electrode and a zinc (II) sulphate (ZnSO_4) electrolyte solution, and a copper half-cell, containing a copper electrode and a copper (II) sulphate solution.

In each half-cell a half-reaction takes place:

At the copper plate, there was an increase in mass. This means that Cu^{2+} ions from the copper (II) sulfate solution were deposited onto the plate as atoms of copper metal. The half-reaction that takes place at the copper plate is:

$$\mathrm{Cu}^{2+}_{\mathrm{(aq)}}\!+\!2\mathrm{e}^{-}
ightarrow\mathrm{Cu}_{\mathrm{(s)}}$$

As electrons are gained by the copper ions this is the reduction half-reaction.

At the zinc plate, there was a decrease in mass. This means that some of the solid zinc goes into solution as Zn^{2+} ions. The electrons remain on the zinc plate, giving it a negative charge. The half-reaction that takes place at the zinc plate is:

$${
m Zn_{(s)}}
ightarrow {
m Zn_{(aq)}^{2+}} {
m + 2e^{-2}}$$

As electrons are lost by the zinc atoms this is the oxidation half-reaction.

You can then combine the two half-reactions from these two half-cells to get the overall reaction: $Zn_{(s)}+Cu^{2+}_{(aq)}+2e^{-} \rightarrow Zn^{2+}_{(aq)}+Cu_{(s)}+2e^{-}$

or, if we cancel the electrons: $Zn_{(s)}{+}Cu^{2+}_{(aq)} \rightarrow Zn^{2+}_{(aq)}{+}Cu_{(s)}$ It is possible to look at the half-reaction taking place in a half-cell and determine which electrode is the anode and which is the cathode.



The silver ions take electrons and are reduced to form solid silver. This makes the silver electrode (cathode) positive.



Exercise 1.4

- 1. In each of the following equations, say which elements in the reactants are oxidised and which are reduced.
 - a. $CuO_{(s)} + H_{2(g)} \rightarrow Cu_{(s)} + H_2O_{(g)}$
 - b. $2NO_{(g)} + 2CO_{(g)} \rightarrow N_{2(g)} + 2CO_{2(g)}$
 - c. $Mg_{(s)}$ +FeSO_{4(aq)} \rightarrow MgSO_{4(aq)}+Fe_(s)
- 2. The following half-reactions take place in an electrochemical cell:

 $\mathrm{Fe}_{(\mathrm{s})}
ightarrow \mathrm{Fe}^{3+}_{(\mathrm{aq})} \! + \! 3\mathrm{e}^{-1}$

 $\mathrm{Fe}_{\mathrm{(aq)}}^{2+}{+}2\mathrm{e}^{-}{
ightarrow}\mathrm{Fe}_{\mathrm{(s)}}$

- a. Which is the oxidation half-reaction?
- b. Which is the reduction half-reaction?
- c. Name the oxidising agent.
- d. Name the reducing agent.
- e. Use standard notation to represent this electrochemical cell.
- 3. For the following cell:

 $\mathrm{Mg}_{\mathrm{(s)}}\left|\mathrm{Mg}_{\mathrm{(aq)}}^{2+}\right|\left|\mathrm{Mn}_{\mathrm{(aq)}}^{2+}\right|\mathrm{Mn}_{\mathrm{(s)}}$

- a. Give the cathode and the anode half-reactions.
- b. Give the overall equation for the electrochemical cell.
- c. What metals could be used for the electrodes in this electrochemical cell?
- d. Suggest two electrolytes for this electrochemical cell.
- e. Draw a simple sketch of the complete cell.

The <u>full solutions</u> are at the end of the unit.

Note

To understand galvanic cells reactions and half cells watch this video by M Champagne, called Galvanic Cells and Activity Series La.

Galvanic Cells and Activity Series La (Duration: 09.59)



Battery types

Primary batteries are non-rechargeable batteries. They include zinc carbon batteries, alkaline batteries, button cell batteries and lithium batteries.

- The zinc carbon battery is labelled as 'all purpose' or 'general purpose' and discharges quickly if used continuously. The zinc chloride cells are longer lasting and are labelled 'heavy duty' or 'super heavy duty'.
- Alkaline batteries can last up to ten times longer than zinc batteries but may cost three to five times more.
- Button cell batteries are small, disc-shaped batteries commonly used in hearing aids, medical devices, watches, calculators, and cameras.
- Lithium batteries can last about twice as long as alkaline batteries but are more expensive. Lithium batteries are labelled as such to distinguish them from other battery types.

Secondary batteries are rechargeable batteries. They are available as freestanding units or as built-in components of rechargeable devices. The free-standing units are expensive but save money in the long run since they can be recharged many times. They are the best for devices that get regular use. The most common types include nickel cadmium (Ni-Cd), sealed lead-acid (Pb), nickel metal hydride (Ni- MH), and lithium ion (Li- Ion).

 Nickel-cadmium is the most common type of rechargeable battery. They may be built into rechargeable appliances or sold as freestanding units. A single nickel cadmium battery can replace about 150 alkaline batteries.



Figure 13: A simplified diagram of a nickel cadmium battery

A nickel-cadmium cell has two plates. The active material of the anode is $Ni(OH)_4$ and the cathode is of cadmium (Cd) when fully charged. The electrolyte is a solution of potassium hydroxide (KOH) with a small addition of lithium hydrate which increases the capacity and life of the battery. When the cell is fully charged, its positive plate is of Ni(OH)4 and its negative plate is of cadmium (Cd).

Discharging: When the cell discharges, the potassium hydroxide (KOH) is dissociated into potassium (K^+) and hydroxyl (OH⁻⁻) ions. The hydroxyl ions go to the cathode and potassium ions go to the anode. The following chemical reaction takes place during discharging.

- $_{\circ}~$ At cathode: $\rm Cd^+$ $+~2\rm OH^- \rightarrow \rm Cd(\rm OH)_2 + 2e^-$
- $_{\circ}~$ At anode: $\mathrm{Ni(OH)_4}$ + $\mathrm{H_2O}$ + $\mathrm{2e^-} \rightarrow ~\mathrm{2OH^-} + ~\mathrm{Ni(OH)_2}$

The anode is converted from $Ni(OH)_4$ to $NI(OH)_2$ and cathode is converted from cadmium (Cd) to cadmium hydroxide $Cd(OH)_2$.

Charging: When the battery is charged, the hydroxyl (OH^{-}) ions move towards the anode, whereas the potassium ions (K^+) move towards the cathode.

Since the voltage produced by a single cell is very low, many cells are connected in series to get the desired voltage output and this arrangement is known as the nickel-cadmium battery.

• Sealed lead-acid batteries are used in some camcorders and cellular phones. They are less expensive, but much heavier than other types of rechargeable batteries.



Lead-acid discharging

Figure 14: A simplified diagram of a lead acid battery

The reactions in a lead acid battery is: Cathode: $PbO_{2(s)} + HSO_{(aq)}^{-4} + 3H_{(aq)}^{+} + 2e^{-} \rightarrow PbSO_{4(s)} + 2H_2O_{(l)}$ Anode (oxidation): $Pb_{(s)} + HSO_{(aq)}^{-4} \rightarrow PbSO_{4(s)} + H_{(aq)}^{+} + 2e^{-}$ Overall: $Pb_{(s)} + PbO_{2(s)} + 2HSO_{(aq)}^{4-} + 2H_{(aq)}^{+} \rightarrow 2PbSO_{4(s)} + 2H_2O_{(l)}$

The diluted sulfuric acid molecules break into two parts when the acid dissolves. It will create positive ions $_{2H^+}$ and negative ions $_{SO^{4-}}$.

Charging: To charge a lead-acid battery we need a charging voltage higher than the nominal battery voltage. To charge a battery of 12 V we need to apply at least 13.5 V DC to the battery terminals for charging.

Discharging: During the discharge of a battery, lead sulfate crystals form at both the negative and positive terminals, as well as the release of electrons due to the change in valence charge of the lead. This formation of lead sulfate uses sulfate from the sulfuric acid electrolyte surrounding the battery.

The charging reaction converts the lead sulfate at the negative electrode to lead. At the positive terminal, the reaction converts the lead to lead oxide. As a by-product of this reaction, hydrogen is evolved. During the first part of the charging cycle, the conversion of lead sulfate to lead and lead oxide

is the dominant reaction. As the charging goes on and most of the lead sulfate is converted to either lead or lead dioxide, the charging current electrolyses the water from the electrolyte and both hydrogen and oxygen gas are evolved, a process known as the 'gassing' of the battery.

- Nickel metal hydride batteries are used in computers, cellular phones, and camcorders. They have a low drain and high-energy capacity.
- Lithium-ion batteries are very expensive, but extremely light and high in energy density. They are used in some cellular phones and notebook computers.

In these batteries ions of lithium travel from negative electrodes to positive electrodes inside the battery through the electrolyte during discharge or when a load is connected, and vice versa (positive to the negative electrode) during charging.

These are very high-quality batteries that have high power density and a very low self-discharge which is very good quality.



Figure 15: The basic structure of lithium-ion batteries

Sacrificial anodes and the protection of metals

Galvanisation is a metal coating process in which a ferrous part is coated with a thin layer of zinc. The zinc coating seals the surface of the part from the environment, preventing oxidation and weathering. Galvanisation helps to extend the life of steel parts by providing a barrier between the steel and the atmosphere, preventing iron oxide from forming on the surface of the steel. Galvanisation also provides superior corrosion resistance to parts exposed to the environment. Galvanisation provides a cost-effective solution for coating steel parts, specifically those that will receive significant environmental exposure over their lifetime.

The primary method of galvanisation is 'hot dip galvanization', which has been in use for over 150 years. Sacrificial anodes are highly active metals that are used to prevent a less active material surface from corroding. Sacrificial anodes are created from a metal alloy with a more negative electrochemical potential than the other metal it will be used to protect. The sacrificial anode will be consumed in place of the metal it is protecting, which is why it is referred to as a 'sacrificial' anode.

When metal surfaces come into contact with electrolytes, they undergo an electrochemical reaction known as corrosion. Corrosion is the process of returning a metal to its natural state as an ore and in this process, causing the metal to disintegrate and its structure to grow weak. These metal surfaces are used all around us — from pipelines to buildings to ships. It is important to ensure that these metals last as long as they can and thus necessitates what is known as cathode protection.

Sacrificial anodes are among several forms of cathode protection. Other forms of cathode protection are plating, galvanisation, and the formation of alloys.

Metal in seawater is one such example with the iron metal coming into contact with electrolytes. Under normal circumstances, the iron metal would react with the electrolytes and begin to corrode, growing weaker in structure and disintegrating. The addition of zinc, a sacrificial anode, would prevent the iron metal from 'corroding'. According to the Table of Standard Reduction Potentials, the standard reduction potential of zinc is about -0.76 v. The standard reduction potential of iron is about -0.44 v. This difference in reduction potential means that zinc would oxidise much faster than iron would. In fact, zinc would oxidise completely before iron would begin to react.

The materials used for sacrificial anodes are either relatively pure active metals, such as zinc or magnesium, or are magnesium or aluminium alloys that have been specifically developed for use as sacrificial anodes. In applications where the anodes are buried, a special backfill material surrounds the anode to ensure that the anode will produce the desired output.

Since the sacrificial anode works by introducing another metal surface with a more negative electronegative and much more anodic surface. The current will flow from the newly introduced anode and the protected metal becomes cathodic creating a galvanic cell. The oxidation reactions are transferred from the metal surface to the galvanic anode and will be sacrificed in favour of the protected metal structure.



Figure 16: A partially corroded sacrificial anode on the hull of a shipSacrificial anodes are normally supplied with lead wires to facilitate their connection to the structure being protected. The lead wires may be attached to the structure by welding or mechanical connections. These should have a low resistance and should be insulated to prevent increased resistance or damage due to corrosion. When anodes with cast-in straps are used, the straps can either be welded directly to the structure or the straps can be used as locations for attachment.

A low resistance mechanically adequate attachment is required for good protection and resistance to mechanical damage. In the process of providing electrons for the cathodic protection of a less active metal the more active metal corrodes. The more active metal (anode) is sacrificed to protect the less active

metal (cathode). The amount of corrosion depends on the metal being used as an anode but is directly proportional to the amount of current supplied.

Sacrificial anodes are used to protect the hulls of ships, water heaters, pipelines, distribution systems, above-ground tanks, underground tanks, and refineries. The anodes in sacrificial anode cathodic protection systems must be periodically inspected and replaced when consumed.



Summary

In this unit you have learnt the following:

- Oxidation is the loss of electrons and reduction is the gain of electrons.
- A redox reaction is one where there is always a change in the oxidation numbers (charge) of the elements that are involved in the reaction.
- It is possible to balance redox equations using the half-reactions that take place within the overall reaction.
- An electrochemical reaction is one where either a chemical reaction produces an electric current, or where an electric current causes a chemical reaction to take place.
- In a galvanic cell a chemical reaction produces a current in the external circuit. An example is the zinccopper cell.
- An electrolytic cell is an electrochemical cell that uses electricity to drive a non-spontaneous reaction. In an electrolytic cell, electrolysis occurs, which is a process of separating elements and compounds using an electric current.
- Cells have a number of components. They consist of two electrodes, which are connected to each other by an external circuit wire.
- In a galvanic cell each electrode is placed in a separate container in an electrolyte solution. The two electrolytes are connected by a salt bridge.
- In an electrolytic cell both electrodes are placed in the same container in an electrolyte solution.
- One of the electrodes is the anode, where oxidation takes place. The cathode is the electrode where reduction takes place.
- In a galvanic cell, the build-up of electrons at the anode sets up a potential difference between the two electrodes, and this causes a current to flow in the external circuit.
- Standard cell notation for a galvanic cell has the anode on the left and the cathode on the right. For example:

 $\left. Zn_{(s)} \left| Zn^{2+}_{(aq)} \right| \left| Cu^{2+}_{(aq)} \right| Cu_{(s)}$

| = a phase boundary (solid/aqueous)

|| = the salt bridge

- \cdot There are different types of batteries: disposable (single use) batteries and rechargeable batteries.
- Sacrificial anodes are more reactive metals used to protect less reactive metals.

Unit 1: Assessment

Suggested time to complete: 30 minutes

- 1. Permanganate (VII) ions (MnO^{-4}) oxidise hydrogen peroxide (H_2O_2) to oxygen gas. The reaction is done in an acid medium. During the reaction, the permanganate (VII) ions are reduced to manganese (II) ions (Mn^{2+}) . Write a balanced equation for the reaction.
- 2. Which one of the substances listed below acts as the oxidising agent in the following reaction? $3SO_{2(g)}+Cr_2O_{2(aq)}^{-7}+2H_{(aq)}^+ \rightarrow 3SO_{2(aq)}^{-4}+2Cr_{(aq)}^{3+}+H_2O_{(l)}$
 - а. H⁺
 - b. Cr^{3+}
 - c. SO_2
 - d. $Cr_2O_2^{-7}$

3. For the following cell:

 $\mathbf{Sn}_{(s)}\left|\mathbf{Sn}_{(aq)}^{2+}\right|\left|\mathbf{Ag}_{(aq)}^{+}\right|\mathbf{Ag}_{(s)}$

- a. Give the cathode and anode half-reactions.
- b. Give the overall equation for the electrochemical cell.
- c. Draw a simple sketch of the complete cell.
- 4. Look at the following half-reactions:

$$\begin{split} & \operatorname{Ca}_{(\mathrm{aq})}^{2+} + 2 e^{-} \rightarrow \operatorname{Ca}_{(\mathrm{s})} \\ & \operatorname{Cl}_{2(\mathrm{g})} + 2 e^{-} \rightarrow 2 \operatorname{Cl}_{(\mathrm{aq})} \\ & \operatorname{Fe}_{(\mathrm{aq})}^{3+} + 3 e^{-} \rightarrow \operatorname{Fe}_{(\mathrm{s})} \\ & \operatorname{I}_{2(\mathrm{s})} + 2 e^{-} \rightarrow 2 \operatorname{I}_{(\mathrm{aq})} \end{split}$$

- a. Which substance is the strongest oxidising agent?
- b. Which substance is the strongest reducing agent?
- 5. A cell is set up. There is a negative charge on the anode.
 - a. What type of cell is this?
 - b. What type of reaction is occurring at the cathode?
 - c. Can the electrodes have the same reaction potential?
- 6. For each of the following, say whether the statement is true or false. If it is false, re-write the statement correctly.
 - a. The anode in an electrolytic cell has a negative charge.
 - b. The reaction $2KClO_{3(s)} \rightarrow 2KCl_{(s)} + 3O_{2(g)}$ is an example of a redox reaction.
 - c. Lead is a stronger oxidising agent than nickel.
- 7. For each of the following questions, choose the one correct answer.
 - a. Which one of the following reactions is a redox reaction?
 - i. $\mathrm{HCl}_{(\mathrm{aq})} + \mathrm{NaOH}_{(\mathrm{aq})} \rightarrow \mathrm{NaCl}_{(\mathrm{aq})} + \mathrm{H}_2\mathrm{O}_{(\mathrm{l})}$
 - ii. $AgNO_{3(s)}\!+\!NaI_{(s)}\rightarrow AgI_{(s)}\!+\!NaNO_{3(s)}$
 - iii. $2FeCl_{3(aq)} + 2H_2O_{(l)} + SO_{2(aq)} \rightarrow H_2SO_{4(aq)} + 2HCl_{(aq)} + 2FeCl_{2(aq)}$
 - $\mathsf{iv.} \quad BaCl_{2(s)} + MgSO_{4(s)} \rightarrow MgCl_{2(s)} + BaSO_{4(s)}$
 - b. Consider the reaction represented by the following equation: $Br_{2(l)}+2I_{(aq)} \rightarrow 2Br_{(aq)}+I_{2(s)}$ Which one of the following statements about this reaction is correct?

Unit 1: Electrochemical Cells | 479

- i. bromine is oxidised.
- ii. bromine acts as a reducing agent.
- iii. the iodide ions are oxidised.
- iv. iodine acts as a reducing agent.
- c. The following equations represent two hypothetical half-reactions:

 $\rm X_2{+}2e^{-} \rightleftharpoons 2X^{-} (E^o{=} +1.09~V)$

and

```
\mathrm{Y}^{+}\!+\!\mathrm{e}\,\rightleftharpoons\,\mathrm{Y}\,\,\left(\mathrm{E}^{\mathrm{o}}\!=\!-2.80\,\,\mathrm{V}
ight)
```

Which one of the following substances from these half-reactions has the greatest tendency to lose electrons?

- i. X⁻
- ii. X_2
- iii. Y
- iv. \mathbf{Y}^+

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.

- a. Oxidation is the loss of electrons by a molecule, atom, or ion.
- b. Reduction is the gain of electrons by a molecule, atom, or ion.
- c. The molecule, atom or ion that is reduced, or the molecule, atom or ion that causes oxidation.
- d. The molecule, atom or ion that is oxidised, or the molecule, atom or ion that causes reduction.

2.

a. The oxidation number of Fe in:

Fe loses two electrons to become $Fe^{2+}.$ The iron species $(Fe_{(s)})$ is oxidised.

b. The oxidation number of O is -2. Therefore the oxidation number of O_3 in Fe_2O_3 is -6 and the oxidation number of Fe_2 in Fe_2O_3 is +6. The oxidation number of Fe in: $Fe_2O_{3(s)} \text{ is } + 3$ $Fe_{(s)} \text{ is } 0$ $Fe^{3+}{}_{(s)}+3e^{-} \rightarrow Fe_{(s)}$ In 3^+ gains three electrons to become Fe(.). The iron species (Fe³⁺, ...) is reduced

 Fe^{3+} gains three electrons to become $Fe_{(s)}.$ The iron species $(Fe^{3+}_{(s)})$ is reduced.

Back to Exercise 1.1

Exercise 1.2

1.

a.
$$14H^+ + 2Mn^{2+} + 5BiO_3^- \rightarrow 2MnO_4^- + 5Bi_3^+ + 7H_2O$$

b. $8H_2O + 2P + 5Cu^{2+} \rightarrow 2H_2PO_4^- + 12H^+ + 5Cu$

c.
$$H_2O + 3NO_2 \rightarrow 2NO_2 + NO + 2H^+$$

2.

a.
$$2\mathrm{Al}+2\mathrm{H}_2\mathrm{O}+2\mathrm{OH}^{-}
ightarrow 2\mathrm{AlO}_2^{-}+3\mathrm{H}_2$$

b.
$$8OH^{-} + 5Pb^{2+} + 2IO_{3}^{-} \rightarrow 5PbO_{2} + I_{2} + 4H_{2}O$$

Back to Exercise 1.2

Exercise 1.3

1.

- a. The copper electrode is the anode, oxidation is a loss of electrons at the anode: Anode half-reaction: $Cu_{(s)} \rightarrow Cu_{(aq)}^{2+} + 2e^{-}$ You are told that the silver electrode is the cathode, reduction is a gain of electrons at the cathode: Cathode half-reaction: $Ag_{(aq)}^{+} + e^{-} \rightarrow Ag_{(s)}$
- b. $\begin{array}{ll} Cu_{(s)}\!+\!2Ag^+_{(aq)} \rightarrow Cu^{2+}_{(aq)}\!+\!2Ag_{(s)}\\ & \text{OR}\\ & Cu_{(s)}\!+\!2AgNO_{3(aq)} \rightarrow Cu(NO_3)_{2(aq)}\!+\!2Ag_{(s)} \end{array}$
- c. $\operatorname{Cu}_{(s)}\left|\operatorname{Cu}_{(aq)}^{2+}\right|\left|\operatorname{Ag}_{(aq)}^{+}\right|\operatorname{Ag}_{(s)}$
- d. Remember that conventional current flow is in the opposite direction to electron flow.



conventional current

- a. Electrolysis
- b. es. Reduction is a gain of electrons. In an electrolytic cell the electrode connected to the negative terminal of the battery is negative. When the metal cations encounter this electrode, they gain electrons and form metal atoms that deposit on the electrode.

c. $Ag^+_{(aq)} + e^- \rightarrow Ag_{(s)}$

Back to Exercise 1.3

Exercise 1.4

٦.

a. The oxidation number of O is -2. O is neither oxidised nor reduced. The oxidation number of Cu in:

 ${
m CuO}_{({
m s})}~{
m is}~+2 \ {
m Cu}_{({
m s})}~{
m is}~0$

 $\mathrm{Cu}_{\mathrm{(s)}}^{2+}{+}2\mathrm{e}^{-}
ightarrow\mathrm{Cu}_{\mathrm{(s)}}$

 ${
m Cu}^{(8)}_{2^+}$ gains two electrons to form Cu. Copper is reduced.

The oxidation number of H in:

 $H2_{(g)}$ is 0

 $H_2O_{(g)}$ is +1

 $m H_{2(g)}
ightarrow 2
m H_{(g)}^+ + 2
m e^-$

 $\rm H_2$ loses two electrons to form $\rm 2H^+.$ Hydrogen is oxidised. Note: Each hydrogen atom in $\rm H_2$ releases one electron, and these two electrons are taken up by a $\rm Cu^{2+}$ ion.

b. The oxidation number of O is -2. O is neither oxidised nor reduced.

The oxidation number of N in: $NO_{(g)}$ is +2

 $N_{2(g)}$ is 0

 $2\mathrm{N}^{2+}_{\mathrm{(g)}}{+}4\mathrm{e}^{-}
ightarrow\mathrm{N}_{2\mathrm{(g)}}$

Two N^{2+} each gain two electrons to become $N_2.$ Nitrogen is reduced. The oxidation number of C in:

 C^{2+} loses two electrons to form $\mathrm{C}^{4+}.$ Carbon is oxidised.

c. The oxidation number of the ion SO_4 is -2 as both a reactant and a product. The oxidation number of O is -2. O is neither oxidised nor reduced. The oxidation number of O_4 in SO_2^{-4} is -8. Therefore the oxidation number of S in SO_2^{-4} is+6. S is neither oxidised nor reduced.

The oxidation number of Mg in: $Mg_{(s)}$ is 0 $MgSO_{4(aq)}$ is +2 $Mg_{(s)} \rightarrow Mg_{(aq)}^{2+}+2e^{-}$ Mg loses two electrons to become Mg^{2+} . Magnesium is oxidised.

The oxidation number of Fe in:

 $FeSO_{4(aq)}$ is +2 $Fe_{(s)} \ is \ 0$ $Fe_{(aq)}^{2+}+2e^{-} \rightarrow Fe_{(s)}$ $Fe^{2+}_{(aq)} \text{ gains two electrons to become Fe. Iron is reduced.}$

- 2.
- a. $\mathrm{Fe}_{\mathrm{(s)}}
 ightarrow \mathrm{Fe}_{\mathrm{(aq)}}^{3+} + 3\mathrm{e}^{-3}$
- b. $\mathrm{Fe}_{\mathrm{(ag)}}^{2+}{+}2\mathrm{e}^{-}{}\rightarrow\mathrm{Fe}_{\mathrm{(s)}}$
- c. An oxidising agent causes oxidation and is itself reduced: $Fe_{(aq)}^{2+}$
- d. A reducing agent causes reduction and is itself oxidised: $Fe_{(s)}$
- e. Oxidation is loss at the anode, therefore $Fe_{(s)}$ in a oxidation half-reaction is the anode. Reduction is gain at the cathode, therefore $Fe_{(s)}$ in a reduction half-reaction is the cathode. $Fe_{(s)} \left| Fe_{(aq)}^{3+} \right| \left| Fe_{(aq)}^{2+} \right| Fe_{(s)}$

3.

a. The cathode is written on the right; therefore manganese is the cathode. Reduction is gain at the cathode.

 ${
m Mn}^{2+}_{(
m ac)}{
m +}2{
m e}^{-}
ightarrow {
m Mn}_{(
m s)}$

The anode is written on the left; therefore magnesium is the anode. Oxidation is loss at the anode. $Mg_{(s)} \rightarrow Mg_{(aq)}^{2+} + 2e^{-2}$

- b. $Mg_{(s)} + Mn_{(aq)}^{2+} \rightarrow Mg_{(aq)}^{2+} + Mn_{(s)}$
- c. Magnesium metal and manganese metal.
- d. Magnesium sulfate and manganese sulfate, or magnesium nitrate and manganese nitrate

e.



Back to Exercise 1.4

Exercise 1.5

1. Sacrificial anodes are used to protect metal structures from corroding.

2. Sacrificial anodes work by oxidising more quickly than the metal it is protecting, being consumed completely before the other metal reacts with the electrolytes.

3. Several different forms of cathode protection are forming alloys, plating, and galvanizing the metal.

Back to Exercise 1.5

Unit 1: Assessment

1. The unbalanced reduction half-reaction is:

 $\mathrm{MnO}_{\mathrm{(aq)}}^{-4}
ightarrow \mathrm{Mn}_{\mathrm{(aq)}}^{2+}$

As this is in an acid medium, we can add water molecules to the right and $\rm H^+$ ions to the left to balance the oxygen and hydrogen atoms:

 ${
m MnO_{(aq)}^{-4}\!+\!8H^+}
ightarrow {
m Mn_{(aq)}^{2+}\!+\!4H_2O_{(l)}}$

Balance the charge by adding five electrons to the left (this makes sense as this is the reduction half-reaction, and $Mn^{7+} \rightarrow Mn^{2+}$):

$${
m MnO_{(aq)}^{-4}\!+\!8H_{(aq)}^{+}\!+\!5e^{-}
ightarrow {
m Mn}_{(aq)}^{2+}\!+\!4{
m H_2O_{(l)}}}$$

The unbalanced oxidation half-reaction is:

$$m H_2O_{2(l)}
ightarrow
m O_{2(g)}$$

Add ${\rm H^+}$ ions to the right to balance the hydrogen atoms:

$${
m H}_2{
m O}_{2({
m l})}
ightarrow {
m O}_{2({
m g})} {+} 2{
m H}^+_{({
m ag})}$$

Balance the charge by adding two electrons to the right (this makes sense as this is the oxidation half-reaction, and $20^{-} \rightarrow 0_2$):

$${
m H_2O_{2(l)}}
ightarrow {
m O_{2(g)}} + 2{
m H_{(aq)}^+} + 2{
m e^-}$$

We multiply the reduction half-reaction by $\underline{2}$ and the oxidation half-reaction by $\underline{5}$ to balance the number of electrons in both equations:

 $2MnO^{\text{-}4}_{(aq)} + 16H^+_{(aq)} + 10\text{e}^{\text{-}} \rightarrow 2Mn^{2+}_{(aq)} + 8H_2O_{(l)}$

 $5 {
m H}_2 {
m O}_{2({
m l})} o 5 {
m O}_{2({
m g})} \!+\! 10 {
m H}^+_{({
m ac})} + 10 {
m e}^-$

Adding the two equations together gives the balanced equation (electrons are equal on both sides and can be removed):

 $2MnO_{(aq)}^{-4} + 16H_{(aq)}^{+} + 5H_2O_{2(l)} \rightarrow 5O_{2(g)} + 10H_{(aq)}^{+} + 2Mn_{(aq)}^{2+} + 8H_2O_{(l)}$

Removing any extra ${\rm H^+}$ ions we get: $2MnO_{(aq)}^{-4}+6H_{(aq)}^++5H_2O_{2(l)}\rightarrow 5O_{2(g)}+2Mn_{(aq)}^{2+}+8H_2O_{(l)}$

- 2. The oxidising agent causes another reactant to be oxidised and is itself reduced.
 - a. H^+ remains as H^+ in H_2O
 - b. Cr^{3+} is a product.
 - c. SO_2 contains S^{4+} , while SO_2^{-4} contains S^{6+} . Therefore S^{4+} loses two electrons and is oxidised.
 - d. $Cr_2O_2^{-7}$ contains Cr^{6+} . Therefore Cr^{6+} gains three electrons to become Cr^{3+} and is reduced. It is the oxidising agent. The answer is d. $Cr_2O_2^{-7}$.

3.

a. The cathode is written on the right; therefore silver is the cathode. Reduction is gain at the cathode.

 $\mathrm{Ag}^+_\mathrm{(aq)}\mathrm{+e^-}
ightarrow\mathrm{Ag}_\mathrm{(s)}$

The anode is written on the left; therefore tin is the anode. Oxidation is loss at the anode. Sn_{(s)} \rightarrow Sn_{(aq)}^{2+}+2e^-

b. To balance the charges the cathode half-reaction must be multiplied by 2. $2Ag^+_{(ac)} + 2e^- \to 2Ag_{(s)}$

 $\mathrm{Sn}_{\mathrm{(s)}}\!+\!2\mathrm{Ag}^+_{\mathrm{(aq)}}
ightarrow\mathrm{Sn}^{2+}_{\mathrm{(aq)}}\!+\!2\mathrm{Ag}_{\mathrm{(s)}}$

c.



4.

- a. Chlorine is the strongest oxidising agent.
- b. Calcium is the strongest reducing agent.
- 5.
- a. Galvanic cell
- b. Reduction
- c. No, a galvanic cell requires two electrodes with different reaction potentials (different metals).

6.

- a. False. The anode in an electrolytic cell has a positive charge.
- b. True
- c. True

7.

- a. Only (iii) has both oxidation and reduction occurring (i.e. a change in the oxidation numbers). iii) $2FeCl_{3(s)} + 2H_2O_{(l)} + SO_{2(aq)} \rightarrow H_2SO_{4(aq)} + 2HCl + 2FeCl_{2(aq)}$
- b. Bromine gains electrons (is reduced), therefore bromine acts as an oxidising agent. Iodide loses electrons (is oxidised), therefore iodide acts as a reducing agent. In the reverse reaction iodine would gain electrons (be reduced) and act as an oxidising agent.
 iii) the iodide ions are oxidised.
- c. Loss of electrons is oxidation. The larger, negative value will be oxidised more easily. $Y \to Y^+ + e^-$ iii) Y

Back to Unit 1: Assessment

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SUBJECT OUTCOME XIII CHEMICAL CHANGE: DESCRIBE, ANALYSE AND APPLY THE RATE AND EXTENT OF CHEMICAL REACTION



Subject outcome

Subject outcome 6.3: Describe, analyse and apply the rate and extent of chemical reaction



Learning outcomes

- Define rates of reactions and identify factors effecting rate and refer to examples.
- State mechanism of reaction and of catalysis.
- Define chemical equilibrium and identify conditions for equilibrium.
- Identify factors affecting equilibrium apply and evaluate the effect on chemical reactions using Le Chatelier's principle.
- Define and interpret equilibrium constant.



Unit 1 outcomes

By the end of this unit you will be able to:

- Define rates of reactions and identify factors effecting rate and refer to examples.
- State mechanism of reaction and of catalysis.
- Define chemical equilibrium and identify conditions for equilibrium.
- Identify factors affecting equilibrium apply and evaluate the effect on chemical reactions using Le Chatelier's principle.
- Define and interpret equilibrium constant.

Unit 1: Rates and extents of chemical reactions

LEIGH KLEYNHANS



Unit outcomes

By the end of this unit you will be able to:

- Define rates of reactions and identify factors effecting rate and refer to examples.
- State mechanism of reaction and of catalysis.
- Define chemical equilibrium and identify conditions for equilibrium.
- Identify factors affecting equilibrium apply and evaluate the effect on chemical reactions using Le Chatelier's principle.
- Define and interpret equilibrium constant.

What you should know

Before you start this unit, make sure you can:

- Describe the kinetic-molecular theory. Refer to <u>level 3 subject outcome 5.4 unit 1</u> if you need help with this.
- Identify, analyse, and apply energy changes in chemical reactions. Refer to <u>level 3 subject outcome 6.1</u> <u>unit 1</u> if you need help with this.
- Write chemical equations and balance them. Refer to <u>level 3 subject outcome 6.2 unit 1</u> if you need help with this.
- Describe the concept of the mole and molar gas volume. Refer to <u>level 3 subject outcome 6.2 unit 2</u> if you need help with this.
- Define the concentration of a substance. Refer to <u>level 2 subject outcome 6.4 unit 1</u> if you need help with this.

Introduction

Parts of the text in this unit were sourced from *Siyavula Physical Science Gr 21 Learner's Book, Chapter 7*, released under a CC-BY licence.

In this unit you will learn how to determine the rate of a reaction and the factors that can affect how fast a reaction happens. You will also learn that some reactions are reversible where the products can be converted back into the reactants and a situation called equilibrium is reached. This is not desirable in the chemical industry, so methods need to be used to manage this process to get maximum yield of the products.

Rate of a reaction

Some chemical reactions are fast, for example, burning a match. Other chemical reactions take a long time,

for example, the rusting of an iron pipe. We will look at why reactions proceed at different rates (speeds) and how we can change the rate of the reaction.

In a chemical reaction, the substances that are undergoing the reaction are called the reactants, while the substances that form as a result of the reaction are called the products. The reaction rate describes how quickly or slowly the reaction takes place. So how do we know whether a reaction is slow or fast? One way of knowing is to look either at how quickly the reactants are used during the reaction or at how quickly the products form. For example, iron and sulfur react according to the following equation:

$$\mathrm{Fe}_{(s)} + \mathrm{S}_{(s)} \
ightarrow \mathrm{FeS}_{(s)}$$

In this reaction, we can observe the speed of the reaction by measuring how long it takes before there is no iron or sulfur left in the reaction vessel. In other words, the reactants have been used up. Alternatively, one could see how quickly the iron sulfide (the product) forms. Since iron sulfide looks very different from either of its reactants, this is easy to do.

Here is another example:

$$2\mathrm{Mg}_{(s)}~+~\mathrm{O}_{2~(g)}~
ightarrow~2\mathrm{MgO}_{(s)}$$

In this case, the reaction rate depends on the speed at which the reactants (oxygen gas and solid magnesium) are used, or the speed at which the product (magnesium oxide) is formed.

Reaction rate is therefore defined as how quickly products are formed, or how quickly reactants are used up during a chemical reaction.

Reaction rates and collision theory

Collision theory is used to explain the rate of a reaction on a microscopic level. You know that all matter is made of particles which are continually moving.

For a reaction to occur, the particles that are reacting must collide with one another. Only a fraction of all the collisions that take place actually cause a chemical change.

These are called *successful* or *effective* collisions. Reactant particles must collide with sufficient energy (called the activation energy) and in the correct orientation at the moment of impact for the reactants to change into products. These *successful collisions* are necessary to break the existing bonds (in the reactants) and form new bonds (in the products).

Factors affecting reaction rate

Any change in conditions under which a reaction is taking place that will increase or decrease the number of successful collisions per unit time will affect the reaction rate. There are a number of factors that can affect the average rate of a reaction. It is important to know these factors so that reaction rates can be controlled. This is particularly important when it comes to industrial reactions, where greater productivity leads to greater profits for companies. The following are some of the factors that affect the average rate of a reaction:

- \cdot the nature of reactants (how reactive the substance is)
- \cdot the concentration of solutions or pressure of gases
- the surface area of solid reactants
- the temperature
- presence of catalysts.



Activity 1.1: Investigate the factors that affect the rate of chemical reactions

Time required: 10 minutes

What you need:

internet connection

What to do:

Watch the video, Factors affecting the rate of a reaction (Demonstrations), on demonstrations of experiments to show how different factors affect the rate of a reaction.

Factors affecting the rate of a reaction (Demonstrations) (Duration: 08.35)



What did you find?

The nature of reactants: the more reactive the reactants, the faster the reaction rate. The concentration of solutions: the higher the concentration of a solution, the faster the reaction rate.

The surface area of solid reactants: increasing the surface area of a solid (by grinding it into a powder) will increase the rate of a reaction.

The temperature: the higher the temperature, the faster the reaction rate.

Adding a catalyst: the presence of a catalyst will increase the rate of a reaction.

1. Concentration

An increase in the concentration implies that there are more particles of solute per unit volume of solution. When more particles are present in a given amount of space, a greater number of collisions will naturally occur between those particles. Since the rate of a reaction is dependent on the number of collisions occurring between reactants, the rate increases as the concentration increases.

Low concentration of red particles



high concentration of red particles

Figure 1: The concentration of red particles has been increased resulting in more collisions per unit time and therefore faster reaction rate

2. Pressure of a gas

When the pressure of a gas is increased, its particles are forced closer together, decreasing the amount of empty space between the particles. Therefore, an increase in the pressure of a gas is also an increase in the concentration of the gas. For gaseous reactions, an increase in pressure increases the rate of reaction as a result of a greater number of collisions between reacting particles.



Figure 2: The pressure has been increased by reducing volume; the particles are closer together resulting in more collisions per unit time

3. Surface area

Surface area is larger when a given amount of a solid is present as smaller pieces. A powdered reactant has a greater surface area than the same mass of reactant as a solid chunk. In order to increase the surface area of a substance, it may be broken into smaller pieces or ground into powder. Increasing the surface area allows more particles to be exposed and therefore increases the number of collisions that can take place per unit time.



red particles in solid block

red particles ground into a powder

Figure 3: A block of solid has been ground into a powder increasing the surface area and thus the number of collisions that can occur per unit time

Serious accidents have been caused by the failure to understand the relationship between surface area and reaction rate. An industry where this is important is flour mills. A grain of wheat is not very flammable. It takes significant effort to get a grain of wheat to burn. When the wheat is ground to make flour, it is ground into a fine powder and some of the powder gets scattered around in the air. A small spark is then sufficient to start a very rapid reaction which can destroy the entire flour mill. Efforts are now made in flour mills to have huge fans circulate the air in the mill through filters to remove the majority of the flour dust particles.

Another example is in the operation of coal mines. Coal in the presence of oxygen burns, but it takes an effort to get the coal to start burning; and once it is burning, it burns slowly because only the surface particles are available to collide with oxygen particles. The inner particles of coal have to wait until the outer surface of the coal lump burns off before they can collide with oxygen. In coal mines, huge blocks of coal must be broken up before the coal can be brought out of the mine. In the process of breaking up the huge blocks of coal, drills are used to drill into the walls of coal. This drilling produces fine coal dust that mixes in the air; a spark from a tool can then cause a massive explosion in the mine. Coal dust explosions have contributed to the death of many miners. In modern coal mines, sprinklers are used to spray water through the air in the mine and this reduces the coal dust in the air, and eliminates coal dust explosions.

4. Temperature

When reactant particles are heated, they move faster and faster. This results in a greater frequency of collisions. A more important effect of the temperature increase is that the collisions occur with a greater force and are thus more likely to surmount the activation energy barrier and go on to form products. Increasing the temperature of a reaction increases the number of effective collisions between reacting particles, so the reaction rate increases.

We use the effects of temperature on reaction rate every day. Food storage is a prime example of how the temperature affects reaction rate. Food is stored in freezers and refrigerators to slow down the processes that cause it to spoil. The decrease in temperature decreases the rate at which food will break down or be broken down by bacteria.

When milk, for example, is stored in the refrigerator, the molecules in the milk have less energy. This means that while molecules will still collide with other molecules, fewer of them will react (which means in this case 'go sour') because the molecules do not have sufficient energy to overcome the activation energy barrier. The molecules do have energy and are colliding, however, and so, over time, even in the refrigerator, the milk will spoil. Eventually the higher energy molecules will gain the energy needed to react and when enough of these reactions occur, the milk becomes 'soured'.

However, if that same carton of milk was at room temperature, the milk would react much more quickly. Most of the molecules would have sufficient energy to overcome the activation energy barrier at room temperature, and many more collisions would occur. This allows for the milk to spoil in a much shorter amount of time.

This is also the reason why most fruits and vegetables ripen in the summer when the temperature is much warmer. When a banana ripens, numerous reactions occur that produce all the compounds that we expect to taste in a banana. But this can only happen if the temperature is high enough to allow the reactions that make those products.

Catalysis

A catalyst increases reaction rates in a slightly different way from other methods of increasing reaction rate. The function of a catalyst is to lower the activation energy so that a greater proportion of the particles have enough energy to react. A catalyst can lower the activation energy for a reaction by:

- orienting the reacting particles in such a way that successful collisions are more likely
- reacting with the reactants to form an intermediate state (called the activated complex) that requires lower energy to form the product.

Some metals, for example platinum, copper and iron, can act as catalysts in certain reactions. In our own bodies, we have enzymes that are catalysts, which help to speed up biological reactions.

Make sure you can illustrate the action of a catalyst using an energy profile diagram. These were studied in the <u>section on energy changes in chemical reactions</u> in level 3.



Figure 4: An energy profile diagram showing a reduced activation energy requirement when a catalyst is present

Much research is done by industrial companies and university research laboratories to find out how catalysts work and to improve their effectiveness. If catalytic activity can be improved, it may be possible to lower the temperature and/or the pressure at which the process occurs and thus save fuel which is one of the major costs in a large-scale chemical process. Further, it may be possible to reduce the amount of reactants that are wasted forming unwanted by-products.

An important application of catalysts is to speed up the hardening of fibreglass resin. Without the addition of a catalyst, the soft resin takes a very long time to harden which would be a disadvantage in the manufacturing industry that uses fibreglass. A similar application of catalysts is used to ensure the rapid hardening of cement.



- d. collision theory
- e. catalyst
- 2. List the six factors that can determine the rate of a reaction.
- 3. State which of the following will happen faster and explain your reasoning:
 - a. lighting a fire using small twigs or large logs
- b. dissolving a lump of sugar in warm water or cold water
- c. taking a headache tablet or a headache powder of the same mass

The <u>full solutions</u> are at the end of the unit.

Chemical equilibrium

To define chemical equilibrium we need to ask some important questions about reactions:

• Does a reaction always run its full course so that all the reactants are used up?

When all the reactants in a reaction are used up the reaction is said to have *gone to completion*. However, in some reactions not all the reactants are used. Not all reactions go to completion.

Reactions that go to completion are *irreversible*. However, in some reactions the reactants form products (in a forward reaction), and the products can change back into reactants (in a reverse reaction).

• Can a reaction reach a point where reactants are still present, but there does not seem to be any further change taking place in the reaction?

In a reversible reaction, the forward reaction starts at a fast rate but as the reactants get used up the rate will slow down. But as the amount of product increases the reverse reaction will start speeding up. Eventually the rate of the forward reaction (*reactants* \rightarrow *products*) equals the rate of the reverse reaction (*products* \rightarrow *reactants*).

At this point there are still reactants present but there does not *appear* to be any further change taking place. As fast as reactants are used up to make products, products are used up to replace reactants. The reaction is said to be in chemical equilibrium.

Chemical equilibrium is the state of a reversible reaction where the rate of the forward reaction equals the rate of the reverse reaction. While a reaction is in equilibrium the concentration of the reactants and products remain constant.



Time required: 5 minutes

What you need:

- 2 glasses
- a saucer
- Vaseline
- water

What to do:

- 1. Half-fill two glasses with water and mark the level of the water in each case.
- 2. Place some Vaseline around the rim of one glass and place the saucer on top of the glass.
- 3. Leave the glasses and, over the course of a day or two, observe how the water levels in the two beakers change.

What did you find?

You should notice that in the beaker that is uncovered, the water level drops more quickly than in the covered beaker. This is because of evaporation. In the beaker that is covered, there is an initial drop in the water level, but after a while evaporation appears to stop and the water level in this beaker is higher than that in the one that is open.



Figure 5: Evaporation and condensation in an open and closed system

In the open glass, liquid water becomes water vapour as a result of evaporation and the water level drops. A small amount of gas molecules will condense again, but because the gas molecules can escape from the system there is much less condensation than evaporation.

In the covered glass, evaporation also takes place. However, in this case, the vapour comes into contact with the surface of the saucer, and it cools and condenses to form liquid water again. This water is returned to the glass. Once condensation has begun, the rate at which the water level drops will start to decrease. At some point, the rate of evaporation will be equal to the rate of condensation, and there will be no change in the water level in the beaker. This can be represented as follows:

liquid ⇒ vapour

In this example, the reaction (in this case, a change in the phase of water) can proceed in either direction. In the forward direction there is a change in phase from liquid to gas (water vapour). The reverse reaction can also take place, when vapour condenses to form liquid again.

An open system is one in which matter or energy can flow into or out of the system. In the liquid-gas demonstration in activity 1.2, the uncovered glass was an example of an open system because the glass could be heated or cooled (a change in energy), and water vapour (the matter) could evaporate from the beaker. An equilibrium will not be established because the water vapour can escape from the glass.

A *closed system* is one in which energy can enter or leave, but matter cannot. The covered glass is an example of a closed system. The glass can still be heated or cooled, but water vapour cannot leave the system because the saucer is a barrier. Condensation changes the vapour to liquid and returns it to the beaker. In other words, there is no loss of matter from the system.



• the reaction must take place in a closed system.

Le Chatelier's principle

Once a reaction reaches equilibrium, the amount of product remains constant. This is not a desirable situation in an industrial process. It is therefore important to find ways to 'shift' the equilibrium. This means changing the conditions under which the reaction is happening to make one of the reactions happen faster than the other. In industry this would be to make the forward reaction happen faster than the reverse reaction so that the yield of the product can be increased.

Any factor that can affect the rate of either the forward or reverse reaction relative to the other can potentially affect the equilibrium position. The following factors can change the chemical equilibrium position of a reaction:

- concentration (for solutions and gases)
- temperature
- pressure (for gaseous reactants).

It is important to understand what effect a change in one of these factors will have on a system that is in chemical equilibrium. However, performing an experiment every time to find out would waste a lot of time. Towards the end of the 1800s the French chemist Henry Louis Le Chatelier came up with principle to predict those effects.

Le Chatelier's principle helps to predict what effect a change in temperature, concentration or pressure will have on the position of the equilibrium in a chemical reaction. This is very important, particularly in industrial applications, where yields must be accurately predicted and maximised.

Le Chatielier's principle states that, when a stress (change in pressure, temperature or concentration) is

applied to a system in equilibrium, the equilibrium will change in such a way as to reduce the effect of the stress.



If the concentration of a substance is changed, the equilibrium will shift to minimise the effect of that change.

If the concentration of a reactant is increased, the equilibrium will shift in the direction of the reaction that *uses* the reactants, so that the reactant concentration decreases. The forward reaction is favoured, or the equilibrium shifts to the right.

The forward reaction is also favoured if the concentration of the product is decreased, so that more product is formed.

If the concentration of a reactant is decreased, the equilibrium will shift in the direction of the reaction that *produces* the reactants, so that the reactant concentration increases. The reverse reaction is favoured, or the equilibrium shifts to the left.

The reverse reaction is also favoured if the concentration of the product is increased, so that product is used.

For example, look at the reaction between sulfur dioxide and oxygen to produce sulfur trioxide:

$$2\mathrm{SO}_2$$
 $_{\mathrm{(g)}}$ + O_2 $_{\mathrm{(g)}}$ \rightleftharpoons $2\mathrm{SO}_3$ $_{\mathrm{(g)}}$

If the SO_2 or O_2 concentration was increased, Le Chatelier's principle predicts that the equilibrium will shift to decrease the concentration of reactants. The forward reaction uses up SO_2 and O_2 , so it will be favoured (happen faster than the reverse) and more SO_3 will be made.

If the SO_2 or O_2 concentration was decreased, Le Chatelier's principle predicts that the equilibrium will shift to increase the concentration of reactants. The reverse reaction makes SO_2 and O_2 , so it will be favoured and the amount of SO_3 will decrease.

If the concentration of SO_3 was decreased, Le Chatelier's principle predicts that the equilibrium will shift to increase the concentration of products. The forward reaction makes SO_3 , so it will be favoured.

If the concentration of SO_3 was increased, Le Chatelier's principle predicts that the equilibrium will shift to decrease the concentration of products. The reverse reaction uses up SO_3 , so it will be favoured.

The effect of temperature on equilibrium

If the temperature of a reaction mixture is changed, the equilibrium will shift to minimise that change.

If the temperature is increased the equilibrium will shift to favour the reaction which will *reduce the temperature*. The endothermic reaction is favoured.

If the temperature is decreased the equilibrium will shift to favour the reaction which will *increase the temperature*. The exothermic reaction is favoured.



 $N_{2~(g)}~+~3H_{2~(g)}\rightleftharpoons~2NH_{3~(g)}~\Delta~H=\text{-92 kJ.mol}^{-1}$

The forward reaction shown is exothermic (shown by the negative value for ΔH). This means that the forward reaction, where nitrogen and hydrogen react to form ammonia, gives off heat, increasing the temperature. In the reverse reaction, where ammonia decomposes into hydrogen and nitrogen gas, heat is taken in by the reaction, cooling the vessel (the reverse reaction is endothermic).

An increase in temperature would therefore favour the reverse reaction because it is endothermic and will take in energy (cooling the container). The yield of ammonia (NH_3) will decrease.

A decrease in temperature will favour the forward reaction because it is exothermic and will release energy (warming the container). The yield of NH_3 will increase.

The effect of pressure on equilibrium

If the pressure of a gaseous reaction mixture is changed the equilibrium will shift to minimise that change.

If the pressure is increased, the equilibrium will shift to favour a decrease in pressure.

If the pressure is decreased, the equilibrium will shift to favour an increase in pressure.

When the volume of a system is increased (and the temperature is constant), the pressure will decrease. There are fewer collisions with the walls of the container and therefore a lower pressure.



Figure 6: The relationship between volume and pressure

The equilibrium will shift in a direction that increases the number of gas molecules so that the pressure is also increased. So, to predict in which direction the equilibrium will shift to change the pressure, you need to look at the ratio of the number of moles of gas in the reactants and number of moles of gas in the products, using the balanced equation for the reaction. (Remember that one mole of any substance contains Avogadro's number of particles)

For example, the equation for the reaction between nitrogen and hydrogen is shown below:

$\mathrm{N_{2~(g)}~+~3H_{2~(g)} \rightleftharpoons~2NH_{3~(g)}}$

The ratio in the balanced equation is 1:3:2. That is, for every 1 mole of N_2 gas, there are 3 moles of H_2 gas and 2 moles of NH_3 gas. Therefore, the ratio is 4 moles of reactant gas to 2 moles of product gas.

An *increase* in pressure will favour the reaction that *decreases* the number of moles of gas. There are fewer moles of product gas than reactant gas, so the *forward reaction* is favoured and the yield of NH_3 will increase.

A decrease in pressure will favour the reaction that *increases* the number of moles of gas. There are more moles of reactant gas, so the *reverse reaction* is favoured and the yield of NH_3 will decrease.



reverse reaction favoured

Figure 7: How changing pressure in a system results in a shift in the equilibrium to counteract that change

The effect of a catalyst on equilibrium

If a catalyst is added to a reaction, both the forward and reverse reaction rates will be increased. If both rates increase, then the concentrations of the reactants and products will remain the same. This means that a catalyst has no effect on the equilibrium position.

However, a catalyst will affect how *quickly equilibrium* is reached. This is very important in industry where the longer a process takes, the more money it costs. So, if a catalyst reduces the amount of time it takes to form specific products, it also reduces the cost of production. Catalysts will therefore be used in conjunction with changing other conditions to favour the forward reaction.



Look at this reaction at equilibrium:

IMAGE/EQUATION

State, and explain, how the position of equilibrium would change if:

- 1. the pressure of the system is increased.
- 2. the temperature of the system is decreased.
- 3. some nitrogen dioxide is removed from the system.

The following strategy should be used to answer the question:

Step 1: Identify the disturbance or stress on the system

Step 2: Use Le Chatelier's principle to decide how the system will respond

Step 3: Look at the given equation and decide whether the rate of the forward reaction or the rate of the reverse reaction is increased to counteract the stress

Step 4: Explain how the chosen reaction counteracts the stress

Step 5: Where appropriate, link equilibrium shift to any observed change in the system, e.g. a colour change, effect on yield

Solutions

- The pressure on the system is increased. According to Le Chatelier's principle the system will respond to reduce the pressure. The reverse reaction will be favoured because it produces fewer moles of gas moles which will decrease the pressure. The colour will go lighter/the yield of NO₂ will decrease.
- 2. The temperature on the system is decreased. According to le Chatelier's principle the system will respond to increase the temperature. The reverse reaction will be favoured as it is exothermic and will release energy to raise the temperature (positive Δ _H value applies to forward reaction). The colour will go lighter/the yield of NO₂ will decrease.
- The concentration of nitrogen dioxide is decreased. According to le Chatelier's principle the system will respond to increase the concentration of nitrogen dioxide. The forward reaction will be favoured as it produced nitrogen dioxide. The colour will go darker/the yield of NO₂ will increase.



Exercise 1.2

1. Bromomethane can be produced using the following chemical reaction:

 $\mathrm{CH}_3\mathrm{OH}_{(g)} + \mathrm{HBr}_{(g)} \rightleftharpoons \mathrm{CH}_3\mathrm{Br}_{(g)} + \mathrm{H}_2\mathrm{O}_{(g)} \Delta\mathrm{H} = -37.2 \mathrm{\ kJ.mol}^{-1}$

For each of the changes given below, predict the effect that the change would have on the yield of bromomethane.

- a. An increase in pressure.
- b. A continuous removal of bromomethane.
- c. An increase in temperature.
- 2. For the reaction below, state five conditions which will increase the yield of ammonia (NH₃). $N_{2(g)} + 3H_{2(g)} \rightleftharpoons 2NH_{3(g)} \Delta H$ is negative

The <u>full solutions</u> are at the end of the unit.

The equilibrium constant

When a reaction is at equilibrium the concentrations (or amounts) of all substances remains constant because as fast as they are being made, they are being used up. It is useful to know how much of each substance is in the container – In particular the amount of products compared to the amount of reactants. A simple ratio from the balanced chemical equation gives us a number called the equilibrium constant (K_c).

 $K_c = \frac{[products]}{[reactants]} \text{ where square brackets mean concentration.}$

The equilibrium constant is defined as the ratio between the concentration of the products and reactants in a chemical reaction.

When the concentration of the reactants is much larger than the concentration of the products, the K_c will be small (less than one). When the concentration of the reactants is less than that of the products, the K_c will be greater than one.

The formula for K_c has the concentration of the products in the numerator and the concentration of reactants in the denominator. A K_c greater than I means that there are more products than reactants in the container at equilibrium.

A K_c between 0 and 1 means that there are more reactants than products in the container at equilibrium.



A high K_{c} value means that:

- the concentration of products is high
- \cdot the reaction has a high yield.

A low K_c value (close to 0) means that:

- the concentration of reactants is high
- the reaction has a low yield.

- 1. Define and give the symbol for the equilibrium constant.
- 2. Why is it useful to know the value of the equilibrium constant for a particular reaction?
- 3. A certain reaction has an equilibrium constant of $7.5 \ge 10^{-2}$. What information does this give about the reaction?

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- The average rate of a reaction describes how quickly reactants are used, or how quickly products form.
- A number of factors can affect the average rate of a reaction. These include the nature of the reactants, the concentration of solutions, the surface area of solids or pressure of gases, the temperature of the reaction and the presence or absence of a catalyst.
- Collision theory provides one way of explaining why each of the factors can affect the average rate of a reaction.
- In order for a reaction to happen the particles of the reactants need to collide with sufficient energy and in the correct orientation. These are called effective or successful collisions.
- A catalyst is used to lower the activation energy so that the reaction is more likely to take place. A catalyst does this by providing an alternative, lower energy pathway for the reaction.
- A catalyst therefore speeds up a reaction but remains unchanged after the reaction is complete.
- A reaction is reversible when reactants can react to form products, and products can react to form the reactants again.
- A reaction is in chemical equilibrium when the rate of the forward reaction equals the rate of the reverse reaction.
- In an open system energy and matter can enter and leave the system. In a closed system energy can enter and leave the system, but matter cannot.
- Le Chatelier's principle states that if an external stress (change in pressure, temperature, or concentration) is applied to a system in chemical equilibrium, the equilibrium will change in such a way as to reduce the effect of the stress.
- The principles of equilibrium are very important in industrial applications so that productivity can be maximised.
- The equilibrium constant (K_c) gives information about the concentration of the reactants and products at equilibrium.
- A high K_c value means that the concentration of products at equilibrium is high, and the reaction has a high yield of the products. A low K_c value means that the concentration of products at equilibrium is low, and the reaction has a low yield of the products.

Unit 1: Assessment

Suggested time to complete: 30 minutes

1. Give two reasons why a collision between reactant particles may not result in the formation of a product.

- 2. Study the descriptions of the following reactions and explain the difference in the rates of reaction for Experiment X and Y in each case:
 - a. Experiment X: a solid piece of magnesium, mass 0.05 g, is placed in 100 ml of hydrochloric acid of concentration 0.01 mol.dm⁻³ and vigorous bubbling occurs.
 Experiment Y: a solid piece of zinc, 0.05 g, is placed in 100 ml of hydrochloric acid of concentration 0.01 mol.dm⁻³ and bubbles form slowly.
 - b. Experiment X: 50 ml of potassium iodate solution of concentration 0.01 mol.dm^{-3} is added to 100 ml of a sodium sulphite and starch mixture. The product makes the solution go blue in 20 s. Experiment Y: 50 ml of potassium iodate solution of concentration $0.001 \text{ mol.dm}^{-3}$ is added to 100 ml of sodium sulphite and starch mixture. The product makes the solution go blue in 35 s.
 - c. Experiment X: A sugar cube is dipped in cigarette ash and ignites as soon as a flame is placed near it.

Experiment Y: An undipped sugar will not ignite even when placed in a flame for $30 \, s$.

- d. Experiment X: Calcium carbonate pieces of total mass 0.1 g are placed in 100 ml of hydrochloric acid of concentration 0.01 mol.dm⁻³ and bubbles form slowly.
 Experiment Y: Calcium carbonate powder of total mass 0.1 g is placed in 100 ml of hydrochloric acid of concentration 0.01 mol.dm⁻³ and bubbles form vigorously.
- 3. Explain why a baker places bread dough in a warm place to rise, rather than in a fridge.
- 4. The following theoretical reaction is at equilibrium. State whether the statements about the reaction are TRUE or FALSE. Correct the statement if it is FALSE.

 $A + B \ \rightleftharpoons \ AB$

- a. The rate of the forward reaction is equal to that of the reverse reaction.
- b. The amount of the reactants and products are equal.
- 5. Define Le Chatelier's principle.
- 6. Water gas is a mixture of carbon monoxide and hydrogen. It is produced by passing steam over redhot coals according to the following reaction:

 $\mathrm{C}_{(s)} + \ \mathrm{H}_2\mathrm{O}_{(g)} \ \rightleftharpoons \ \mathrm{H}_2 \ \mathrm{_{(g)}} \ + \ \mathrm{CO}_{(g)} \ \Delta\mathrm{H}{=}131 \ \mathrm{kJ.mol}^{-1}$

How is the equilibrium affected by:

- a. continually pumping in steam?
- b. using red-hot coals?
- c. continually adding more coal?
- d. operating the system at a high pressure?
- 7. Given the following reaction at equilibrium: $CH_{3(g)} + 2O_{2(g)} \rightleftharpoons CO_{2(g)} + 2H_2O_{(l)} + heat$ use Le Chatelier's principle to explain the effect if the following changes are made.
 - a. more oxygen is added to the equilibrium mixture
 - b. the temperature of the system is decreased
 - c. the pressure of the system is increased.
- 8. In a closed bottle of fizzy cooldrink that is ice cold, the carbon dioxide in the drink is in equilibrium with water as follows:

 $\mathrm{CO}_{2~(g)}\,+\,\mathrm{H}_{2}\mathrm{O}_{(l)}\,\,\rightleftharpoons\,\,\mathrm{H}_{2}\mathrm{CO}_{3~(aq)}\,\,\Delta\mathrm{H}$ is negative

Use Le Chatelier's principle to explain the following:

- a. As soon as the bottle is opened the gas bubbles up and escapes.
- b. If a warm bottle is opened the gas bubbles up more quickly.
- 9. The following reaction is at equilibrium:

IMAGE/EQUATION

- a. Explain what would happen to the colour of the solution if more ${\rm Fe}^{3+}$ was added.
- b. When the solution is heated, the red colour fades. Is the forward reaction exothermic or

endothermic? Explain your answer.

- 10. At a certain temperature, the K_c for the following reaction is 0.18. $PCl_{3 (g)} + Cl_{2 (g)} \rightleftharpoons PCl_{5 (g)} \Delta H$ is negative
 - a. What information does the K_c value indicate about the yield of PCl_5 ?
 - b. List five ways to increase the yield of PCl_5 .

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.

- a. reaction rate: how quickly reactants are used, or how quickly products form.
- b. activation energy: the minimum amount of energy required for a reaction to take place.
- c. concentration: a measure of the number of particles in a given amount of space
- d. collision theory: the theory that explains that a chemical reaction will only take place if particles collide with sufficient energy and in the correct orientation.
- e. catalyst: a substance that can increase the rate of a reaction by lower the amount of activation energy required
- nature of the substances surface area of a solid concentration of a solution pressure of a gas temperature presence of a catalyst
- Lighting a fire using small twigs greater surface area will increase the probability of successful collision per unit time, therefore increase the reaction rate.
 Dissolving a lump of sugar in warm water – a higher temperature means particles will collide with

greater energy (force) and the number of successful collisions per unit time will increase, therefore the reaction rate will increase.

Taking a headache powder – increased surface area will increase the probability of successful collisions per unit time and therefore increase the reaction rate.

Back to Exercise 1.1

Exercise 1.2

1.

- a. No effect the number of moles of gas in the reactants is the same as the number of moles of gas in the products.
- b. The yield of bromomethane would increase if concentration is decreased, the reaction that makes that substance will be favoured. The forward reaction makes bromomethane, therefore the yield will increase.
- c. The yield of bromomethane would decrease if temperature is increased, the endothermic reaction will be favoured to reduce the temperature. The reverse reaction is endothermic (negative Δ H value indicates forward reaction is exothermic).

- 2. To increase yield, the forward reaction needs to be favoured. Ways to do this will be:
 - \cdot increase concentration of N_2 (forward reaction uses up N_2)
 - \cdot increase concentration of H_2 (forward reaction uses up H_2)
 - \cdot decrease concentration of NH_3 (forward reaction makes NH_3)
 - decrease the temperature (forward reaction is exothermic so will release heat to increase temperature)
 - increase the pressure (the products contain more moles of gas compared to the reactants)

Back to Exercise 1.2

Exercise 1.3

- 1. The equilibrium constant is the ratio between the concentration of the products and reactants in a chemical reaction. The symbol is K_c .
- 2. The equilibrium constant indicates whether there are more products in the reaction vessel at equilibrium or more reactants.
- 3. The value for the equilibrium constant is less than1, therefore, when the reaction has reached equilibrium there are more reactants in the container than products. The equilibrium lies to the left and the yield of the product is low.

Back to Exercise 1.3

Unit 1: Assessment

- 1. The colliding particles may not have sufficient energy to overcome the activation energy requirement, or they may not be colliding in the correct orientation in order to break their bonds.
- 2.
- a. Experiment X happened at a faster rate because magnesium is a more reactive metal than zinc, so the nature of the substance determined which reaction happened faster.
- b. Experiment X happened faster because the concentration of the potassium iodate solution was greater. The higher the concentration of a solution, the greater the probability of effective collisions per unit time, therefore the faster the reaction rate.
- c. Experiment X happened faster because the cigarette ash acted as a catalyst. Catalysts speed up reaction rate by lowering the activation energy requirement for the reaction.
- d. Experiment Y happened faster because the magnesium powder has a greater surface area therefore more effective collisions can occur per unit time and speed up the reaction rate.
- 3. Reaction rates are faster at higher temperatures. The reaction in the bread dough that causes it to rise will happen faster in a warm place than in a cold place.

4.

- a. TRUE (definition of equilibrium).
- b. FALSE. The amount of the reactants and products remains constant.
- 5. Le Chatelier's principle states that if an external stress (change in pressure, temperature, or concentration) is applied to a system in chemical equilibrium, the equilibrium will change in such a way as to reduce the effect of the stress.

6.

- a. continually pumping in steam favours the forward reaction (the equilibrium shifts to the right), and the concentration of the products will increase.
- b. using red-hot coals favours the forward reaction because it is endothermic and will absorb the

excess heat. The equilibrium shifts to the right and the concentration of the products will increase.

- c. continually adding more coal will not affect the equilibrium. It is not possible to change the concentration of a solid only solutions and gases.
- d. high pressure will favour the reverse reaction because the reactants contain less moles of gas than the products.

7.

- a. If more oxygen is added to the system the equilibrium will shift to use up oxygen. The forward reaction will therefore be favoured, and the concentration of the products will increase.
- b. If the temperature is decreased the equilibrium will shift to produce heat. The forward reaction produces heat so it will be favoured, and the concentration of the products will increase.
- c. If the pressure is increased the reaction that produces fewer moles of gas will be favoured in order to reduce the pressure. The forward reaction will be favoured because it produces only one mole of gas compared to the reverse reaction which produces three moles of gas.

8.

- a. Opening the bottle reduces the pressure and this favours the reverse reaction in order to produce more moles of gas in order to increase the pressure, but the bottle is open so the gas escapes.
- b. Heating favours the reverse reaction as it is endothermic. The reverse reaction produces the carbon dioxide gas.

9.

- a. The solution would become darker red because increasing the concentration of ${\rm Fe}^{3+}$ will favour the forward reaction in order to use up the extra ${\rm Fe}^{3+}$ and the forward reaction produces the red ${\rm FeCNS}^{2+}$ ions.
- b. If the red colour fades, the reverse reaction is being favoured by heating. Therefore the reverse reaction must be endothermic. Thus the forward reaction must be exothermic.

10.

- a. The yield of PCl_3 is low.
- b. The forward reaction must be favoured so the following changes can be made:
 - \cdot increase the concentration of PCl_3
 - \cdot increase the concentration of Cl_2
 - \cdot decrease the concentration of PCl_5
 - decrease the temperature (forward reaction is exothermic)
 - decrease the pressure (forward reaction produces less moles of gas).

Back to Unit 1: Assessment

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SUBJECT OUTCOME XIV CHEMICAL SYSTEMS AND INDUSTRY: IDENTIFY AND CRITICALLY EVALUATE THE IMPACT OF EXPLOITING THE LITHOSPHERE ON THE QUALITY OF HUMAN, ENVIRONMENTAL, AND SOCIO-ECONOMIC DEVELOPMENT

Subject outcome

Subject outcome 7.1: Identify and critically evaluate the impact of exploiting the lithosphere on the quality of human, environmental, and socio-economic development



(*)

Learning outcomes

· Identify and evaluate the energy resources, their use, pollution, and renewability.

Range: Coal, oil, gas, solar, hydro, wind, electrochemical, nuclear, geothermal, vegetative, and other energy sources.



Unit 1 outcomes

By the end of this unit you will be able to:

• Identify and evaluate renewable and non-renewable energy resources in terms of their use and ability to cause pollution.

Unit 1: The Lithosphere

EMMA HARRAGE



What you should know

Before you start this unit, make sure you can:

- Understand air pollution. Refer to level 3 subject outcome 7.1 unit 2 if you need help with this.
- Understand water pollution. Refer to level 2 subject outcome 7.1 unit 2 if you need help with this.
- Understand how an electrochemical cell works. Refer to <u>level 4 subject outcome 6.2 unit 1</u> if you need help with this.

Introduction

The lithosphere is the upper crust of the earth's surface where most life on earth can be found. Within the lithosphere, in various parts of the world, fossil fuels can be found. These are oil, coal, and gas. Most of the energy used in daily life comes from burning fossil fuels. South Africa has large coal reserves and burns coal to generate electricity. There is a growing demand for 'cleaner' energy because of the pollutants released when fossil fuels are burned and their contribution to global warming and climate change. Alternative, renewable sources of energy generation by wind, solar, hydro, and geothermal are increasingly being used around the world because they do not cause pollution which contributes to climate change.

Renewable and non-renewable resources

A natural resource is something supplied by nature that helps support life. When you think of natural resources, you may think of minerals and fossil fuels. However, ecosystems and the services they provide are also natural resources. Biodiversity is a natural resource as well.

- Renewable resources can be replenished by natural processes as quickly as humans use them.
 Examples include sunlight and wind. They are in no danger of being used up (see figure 1 below).
 Metals and other minerals are renewable too. They are not destroyed when they are used and can be recycled.
- Non-renewable resources are natural resources that exist in fixed amounts and can be used up.
 Examples include fossil fuels such as petroleum, coal, and natural gas. These fuels formed from the remains of plants over hundreds of millions of years. We are using them up far faster than they could ever be replaced. At current rates of use, petroleum will be used up in just a few decades and coal in less than 300 years. Nuclear power is also considered to be a non-renewable resource because it uses up uranium, which will sooner or later run out. It also produces harmful wastes that are difficult to

dispose of safely.



Figure 1: The different types of energy resources available; the images circled in red are non-renewable and the images in green are renewable sources

Non-renewable sources of energy

Fossil fuels are a chemical store of energy and include coal, oil, and natural gas. Most of South Africa's electricity is produced by burning fossil fuels, mainly coal (90%).

Oil

As we learnt in level 4 subject outcome 5.4 unit 4, [link to: L4 SO 5.4 Unit 4] crude oil is separated by fractional distillation and used to make the fuels petrol and diesel for cars, trucks, and motorbikes. It can be used to generate electricity, although its use as an energy source is in decline.



https://howmuch.net/articles/worlds-biggest-crude-oil-reserves-by-country Central Intelligence Agency - https://www.cia.gov/library

Figure 2: Global oil reserves; South Africa has oil reserves of $15\,\,000\,\,000$ barrels



Figure 3: Countries such as Nigeria and the DRC use oil to generate electricity

Three technologies are used to convert oil into electricity:

- Conventional steam Oil is burned to heat water to create steam to generate electricity.
- Combustion turbine Oil is burned under pressure to produce hot exhaust gases which spin a turbine to generate electricity.
- Combined-cycle technology Oil is first combusted in a combustion turbine, using the heated exhaust gases to generate electricity. After these exhaust gases are recovered, they heat water in a boiler, creating steam to drive a second turbine.

Environmental impacts

Burning oil for electricity pollutes the air, water, and land but some of the worst environmental woes associated with oil are linked to drilling, transporting, and refining.

Burning oil to generate electricity produces significant air pollution in the forms of nitrogen oxides, and, depending on the sulfur content of the oil, sulfur dioxide and particulates. Carbon dioxide and methane (as well as other greenhouse gases), heavy metals such as mercury, and volatile organic compounds (which contribute to ground-level ozone) can be released into the atmosphere from the smokestack of an oil-burning power plant.

The operation of oil-fired power plants also impacts water, land use and solid waste disposal. Like the operations of other conventional steam technologies, oil-fired conventional steam plants require large amounts of water for steam and cooling and can negatively impact local water resources and aquatic habitats. Sludge and oil residue that are not consumed during combustion become solid waste and contain toxic and hazardous materials.

Drilling also produces a long list of air pollutants, toxic and hazardous materials, and emissions of hydrogen

sulfide, a highly flammable and toxic gas. All these emissions can impact the health and safety of workers and wildlife. Loss of huge stretches of wildlife habitat also occur during drilling. Refineries, too, spew pollution into the air, water, and land (in the form of hazardous wastes). Oil transportation accidents can result in catastrophic damage killing thousands of fish, birds, other wildlife, plants, and the contamination of soil.



Crude oil is a non-renewable energy source because it takes millions of years to produce crude oil and so we cannot produce more when the existing reserves are finished.

Coal

Coal has traditionally dominated the energy supply sector in South Africa, from as early as 1880 when coal from the Vereeniging area was supplied to the Kimberly diamond fields.



Figure 4: Internationally, coal is currently the most widely used primary fuel, accounting for approximately 36% of the world's electricity production

South Africa produces an average of 224 million tons of coal annually, making it the fifth largest coal producing country in the world. South Africa's coal reserves are estimated at 53 billion tonnes, and with our present production rate there should be almost 200 years of coal supply left.

Producing electricity from coal starts when the coal is pulverised in huge mills into a fine powder before it is blown into huge kettles, called boilers. Due to the heat in the boiler, the coal particles combust and burn to generate heat to turn water into steam. The steam from the boilers is used to turn the blades of a giant fan or propeller, called a turbine. The turbine turns a coil made of copper wire (the rotor) inside a magnet (the stator). Together they make up the generator. The generator produces an electric current, which is sent to the homes and factories of consumers via power lines.

The advantages of coal

Coal has several advantages when it comes to producing energy.

- 1. Reliability. One of the greatest advantages of coal fired power plants is reliability. Coal's ability to supply power during peak power demand either as base power or as off-peak power is greatly valued as a power plant fuel. Advanced pulverised coal fired power plants are designed using this feature to support the grid system in avoiding blackouts.
- 2. Affordability. Energy produced from coal fired plants is cheaper and more affordable than other energy sources. Since coal is abundant, it is cheap to produce power using this fuel. Moreover, it is not expensive to extract and mine from coal deposits. Consequently, its price remains low compared to other fuel and energy sources.
- 3. Abundance. There are approximately over 300 years of economic coal deposits still accessible globally. With this great amount of coal available for use, coal fired plants can be continuously fuelled for many years to come.
- 4. Known technologies. The production and use of coal as a fuel are well understood, and the technology required in producing it is constantly advancing.
- 5. Safety. Generally, coal fired plants are considered safer than nuclear power plants. A coal power plant's failure is certainly not likely to cause catastrophic events such as a nuclear meltdown would. Additionally, the welfare and productivity of coal industry employees has greatly improved over the years.

The disadvantages of coal

On the other hand, there are also some significant disadvantages of coal fired plants including greenhouse gas (GHG) emissions, mining destruction, generation of millions of tons of waste, and the emission of harmful substances.

- 1. Greenhouse gas emissions. The by-products of combustion cause a lot of pollution and contribute to global warming. The increased carbon emissions by coal fired plants has led to further global warming which results in climate change.
- 2. Mining destruction. Mining of coal not only results in the destruction of habitat and scenery, but it also displaces humans as well. In many countries where coal is actively mined, many people are displaced in huge numbers due to the pitting of the earth brought about by underground mining. Places near coal mines are unsafe for human habitation as the land could cave in at any time.
- 3. Generation of millions of tons of waste. These waste products contribute to waste disposal problems and also contain harmful substances.
- 4. Emission of harmful substances. These include mercury, sulfur dioxide, carbon monoxide, mercury, selenium, and arsenic. These harmful substances not only cause acid rain but also are very harmful to humans as well.

Gas

Natural gas power plants generate electricity by burning natural gas as their fuel. There are many types of natural gas power plants which all generate electricity but serve different purposes. All natural gas plants use a gas turbine; natural gas is added, along with a stream of air, which combusts and expands through this turbine causing a generator to spin a magnet, making electricity.

South Africa currently generates 5% of its total electricity using gas from seven plants located mainly in the Western Cape. The newest plant is the 140 MW plant located at Sasolburg.

South Africa has no natural gas reserves and all the natural gas used is imported.



Figure 5: A natural gas power plant

Natural gas power plants are cheap and quick to build. They also have very high thermodynamic efficiencies compared to other power plants. Burning of natural gas produces fewer pollutants like NO_x , SO_x and particulate matter than coal and oil.

On the other hand, natural gas plants have significantly higher emissions than a nuclear power plant, but air quality tends to improve (i.e. reduction in smog) when switching to natural gas plants from coal plants.

Natural gas plants significantly contribute to climate change, and that contribution is growing. Natural gas power plants produce considerable carbon dioxide, although less than coal plants do. Also, the process of getting natural gas from where it is mined to the power plants leads to a considerable release of methane (natural gas that leaks into the atmosphere). If natural gas plants are used to produce electricity their emissions will continue to warm the planet in dangerous ways.

Nuclear power

South Africa is among the top countries in the world with uranium reserves, and accounts for a significant reserve base of an estimated $279\ 100\ {\rm tonnes}$ of uranium, or around 5.2 of proven global reserves.

Uranium is used in a nuclear power station to produce energy, which contributes about 3% of the total energy supply in South Africa. All the uranium used at the Koeberg Power Station is produced locally. As mentioned in <u>level 3 subject outcome 7.1 unit 2</u>, South Africa has one nuclear power plant located in the Western Cape. A new plant located in Thyspunt in the Eastern Cape has been approved as of September 2021.

The disadvantages of nuclear energy

There are however, several disadvantages of using nuclear power.

- 1. Environmental impact: The process of mining and refining uranium is not clean. Transporting nuclear fuel to and from plants involves a pollution hazard. Also, once the fuel is used, you can't simply take it to the landfill it is radioactive and dangerous.
- 2. Radioactive waste disposal: As a rule, a nuclear power plant creates 20 metric tons of nuclear fuel per year, and a large amount of nuclear waste. The biggest part of this waste emits radiation and heat, meaning that it will decay any compartment that holds it. It can also cause damage to living things in and around the plants.

Nuclear power plants create a lot of low-level radioactive waste. Over time, used nuclear fuel decays to safe radioactive levels, however, this takes a countless number of years. Even low-level radioactive waste takes hundreds of years to achieve adequate levels of safety. South Africa's radioactive waste is currently stored in the Kalahari Desert.

3. Nuclear accidents: The worst nuclear accident to date was the Chernobyl disaster which occurred in 1986 in Ukraine. The accident killed 31 people directly and damaged approximately 7 billion dollars' worth of property. Radioactive fallout from the accident was concentrated in areas of Belarus, Ukraine, and Russia. Studies have estimated as many as over a million eventual cancer deaths from Chernobyl. To date, Chernobyl remains a closed off area.

The most recent nuclear disaster was the Fukushima Daiichi nuclear disaster in 2011 in Japan. The cause of the disaster was an earthquake and tsunami. It was the most severe nuclear accident since the Chernobyl disaster.

On detecting the earthquake, the active reactors automatically shut down their normal powergenerating fission reactions. Because of these shutdowns and other electrical grid supply problems, the reactors' electricity supply failed, and their emergency diesel generators automatically started. However, the earthquake had also generated a tsunami 14 metres high that arrived shortly afterwards and swept over the plant's seawall and then flooded the lower parts of the reactors. This flooding caused the failure of the emergency generators and loss of power to the circulating pumps. The resultant loss of reactor core cooling led to three nuclear meltdowns, three hydrogen explosions, and the release of radioactive contamination.

The government declared an evacuation zone around the plant, with a 20 km radius, and almost 200 000 residents were evacuated, large amounts of water contaminated with radioactive isotopes were released into the Pacific Ocean during and after the disaster.

- 4. High cost: The initial costs for building a nuclear power plant are steep. South Africa scrapped plans to add 9.6 GW of nuclear power to its energy mix due to the cost, which was estimated anywhere between \$34 84 billion (approximately R476 billion R1.76 trillion). The nuclear plants are cheap to run and produce inexpensive fuel, but the initial costs are huge.
- 5. Uranium is finite: Unlike fossil fuels that are available to most of the countries, uranium is a very scarce resource and exists in only a few countries. Typical renewable energy sources such as solar and wind are in infinite supply. Uranium is in limited supply and, although currently abundant, there is still the risk of running out eventually.

The advantages of nuclear energy

There still remain some advantages of using nuclear power.

 Low pollution: Nuclear power creates fewer greenhouse emissions. It has been determined that the number of greenhouse gases have decreased by almost half because of the use of nuclear power. This avoids more than 470 million metric tons of carbon each year, which is the equivalent of removing 100 million cars off the road. Nuclear energy has the least effect on nature since it doesn't discharge any gasses such as methane and carbon dioxide, which are the primary 'greenhouse gasses'.

Nuclear advocacy group the World Nuclear Association found that the average emissions for nuclear are 29 tonnes of CO_2 per gigawatt-hour (GWh) of energy produced. This compares favourably with

renewable sources like solar and wind and even more favourably with fossil fuels like coal (888 tonnes per GWh).

- 2. High power output: It has the capacity to meet city and industrial needs with just one reactor. A relatively small amount of uranium can be used to fuel a 1 000 megawatt electric plant and provide enough electricity to power a city of about half a million people.
- 3. Stable base load energy: This means that it can be used in conjunction with other forms of renewable energy. For example, when the wind is blowing, nuclear plants can adjust energy output to be lower. Conversely, when the wind is not blowing, and greater energy is needed, nuclear energy can be adjusted to compensate for the lack of wind (or solar) generated power.
- 4. Low operating costs: Nuclear power produces very inexpensive electricity and is cheaper than gas, coal, or any other fossil fuel plants. The cost of the uranium is low.
- 5. Reliability: It is estimated that with the current rate of consumption of uranium, we have enough uranium for another 70 80 years. As solar and wind energy are dependent upon weather conditions, the nuclear power plant has no such constraints and can run without disruption in any climatic condition.
- 6. More proficient than fossil fuels: The amount of fuel required by the nuclear power plant is comparatively less than what is required by other power plants, as the energy released by nuclear fission is approximately ten million times greater than the amount of energy released by fossil fuels per atom.
- 7. Not reliant on fossil fuels: Coal and natural gas power plants discharge carbon dioxide into the air, which causes several environmental issues. With nuclear power plants, carbon emissions are insignificant.
- 8. Potential for an alternative: Nuclear energy is not a renewable resource. Uranium is limited and cannot be produced again and again on demand. Thorium is a alternative that has recently been identified as a nuclear fuel. China, Russia, and India already have plans to start using thorium to fuel their reactors soon.
- 9. Economic impact: Nuclear power provides many benefits to the economy with the number of jobs and the prosperity a new plant brings.



Exercise 1.1

- 1. Give one advantage and one disadvantage of using nuclear energy to generate electricity.
- 2. List two disadvantages, excluding acid rain, of a heavy reliance on fossil fuels to produce electricity.
- 3. Give three reasons why South Africa is dependent on coal to generate most of its electricity.

The <u>full solutions</u> are at the end of the unit.

Renewable sources of energy

South Africa is regarded as a prime candidate for increased use of renewable energy with its abundant natural resources of sun and wind. The country is highly dependent on burning coal for power generation at the moment, but at the same time South Africa has an abundance of sunshine which lends itself very well to solar water heating and electricity generation.

Given the ever-rising cost of traditional fossil fuels-based energy, renewable energy is becoming a viable option. The Renewable Energy Independent Power Producers Procurement Programme (REIPPPP) has to date, attracted investment to the value of R209.7 billion.



Figure 6: Map of the current different renewable energy plants in South Africa



SA utility scale RE technologies by province

Figure 7: The graph shows the cumulative number of megawatts procured per technology and is grouped by province

The current new capacity projections include 14 725 MW of renewable energy, comprising of solar PV (6 225 MW), wind (6 360 MW), concentrated solar power (CSP) (1 200 MW), small hydro (195 MW), biomass (210 MW), biogas (110 MW).

Wind farms in South Africa

Winds are caused by the rotation of the earth, heating of the atmosphere by the sun, and the earth's surface irregularities. We can harness wind energy and use it to generate power if the sun shines and the wind blows.

Wind turbines are a relatively simple system of generating electricity and have not changed much over the centuries. As the wind rotates the blades – which face into the wind and are tilted to generate the greatest rotation – the blades rotate a shaft and a generator, which turn the energy into electricity. Modern wind turbines will also have sensors which detect the direction and power of the wind, so they can be rotated toward the wind or shut down if the wind is too low, or too powerful.

Crucially, electrical energy is lost if it is transported great distances, so the closer the wind farms are to the grid, the more efficient they become.

South Africa has thirty-six winds farms which are fully operational or under construction. The majority are in the Eastern Cape.

These projects form part of the REIPPP programme, which aims to procure 17 800 MW of renewable energy capacity by 2030.

Wind is a renewable resource with wind farms having a virtually non-existent carbon footprint once the turbines have been constructed.

The advantages of wind farms

Wind farms have a number of advantages:

- 1. Wind energy is a clean source of power: Wind does not pollute the air or require any destructive chemicals. As a result, wind energy lessens our reliance on fossil fuels.
- 2. Renewable source: The wind is limitless.
- 3. Wind energy has low operating costs: The wind farms or individual turbines can be expensive to install. However, once installed and running, operating costs are relatively low; fuel (wind) is free, and the turbines don't require too much maintenance over the course of their life.
- 4. Prices are decreasing: Prices of turbines have decreased by over 80% since 1980. Thanks to technological advancements and increased demand, prices are expected to keep decreasing in the foreseeable future.
- 5. Wind farms can be built on existing farms: Wind turbines are incredibly space-efficient and can be installed on existing farms or agricultural land in rural areas where they can be a source of income for the farmers as wind plant owners make payments to farmers for the use of their land for electricity generation.
- 6. Conserves and keeps water clean: Turbines produce no particulate emissions that contribute to mercury contamination in our lakes and streams. Wind energy also conserves water resources. For producing the same amount of electricity, nuclear power takes about 600 times more water than wind, and coal takes about 500 times more water than wind.

The disadvantages of wind energy

Wind farms also have a number of disadvantages:

- 1. Wind reliability: Wind does not generally blow reliably.
- 2. Wind turbines could be a threat to wildlife: The edges of wind turbines can be unsafe for wildlife, especially birds and other flying creatures that may be in the area.
- 3. Wind turbines could lead to noise and visual pollution.
- 4. Expensive to set up: The manufacturing and installation of wind turbines require heavy upfront investments.
- 5. Wind power can be harnessed at certain locations only: Wind energy can only be harnessed at certain locations where the speed of the wind is high. Since they are mostly set up in remote areas, transmission lines must be built to bring the power generated to cities, which requires extra investment to set up the infrastructure.
- 6. Effect on the environment: Wind farms need wide open areas to set up wind turbines and some believe they are an eyesore.

Solar power

Solar power in South Africa includes photovoltaics (PV) as well as concentrated solar power (CSP). In 2016, South Africa had 1 329 MW of installed solar power capacity. Installed capacity is expected to reach 8 400 MW by 2030. South Africa currently has 55 solar farms, and an increasing number of private individuals are installing solar panels on the roofs of their houses and businesses.

South Africa is fortunate in the amount of sunlight it receives every day. Solar energy can be regarded as one of the cleanest and greenest energy sources that is sufficiently developed to deliver significant amounts of electricity. The largest solar farm located in De Aar is the largest solar power plant located in the Southern hemisphere and the Middle East. The entire solar farm facility is spread over a surface of almost 100 hectares and consists of 700 000 solar panels, delivering 50 MW. On very sunny days, the power plant produces much more power than the grid demands, so the excess energy goes to waste as it cannot be stored.



Figure 8: The Northern Cape and the Eastern Cape are both sun-drenched regions boasting some of the best conditions for solar power in the world

South Africa's solar farms employ one of two methods to harness the sun's energy:

 Photovoltaics (Solar PV) is a form of active solar technology that was discovered in 1839 by 19-year-old French physicist Alexandre-Edmond Becquerel. Becquerel discovered that when he placed silverchloride in an acidic solution and exposed it to sunlight, the platinum electrodes attached to it generated an electric current. This process of generating electricity directly from solar radiation is called the photovoltaic effect, or photovoltaics.

Photovoltaic arrays usually involve solar panels, a collection of dozens or even hundreds of solar cells. Each solar cell contains a semiconductor, usually made of silicon. When the semiconductor absorbs sunlight, electrons become free to move. An electrical field directs these free electrons forming an electric current, flowing in one direction. Metal contacts at the top and bottom of a solar cell direct that current to an external object. The external object can be as small as a solar-powered calculator or as large as a power station.

Photovoltaic technology can also be installed on a smaller scale. Solar panels and cells can be fixed to the roofs or exterior walls of buildings, supplying electricity for the structure.



Figure 9: Jasper solar farm, made up of 25 000 photovoltaic modules, is outside Kimberley in the Northern Cape

2. Concentrated Solar Power (CSP). CSP technology uses lenses and mirrors to concentrate sunlight from a large area into a much smaller area. This intense area of radiation heats a fluid, which in turn generates electricity or fuels another process.

Solar power towers use heliostats, flat mirrors that turn to follow the sun's arc through the sky. The mirrors are arranged around a central 'collector tower', and reflect sunlight into a concentrated ray of light that shines on a focal point on the tower.



Figure 10: A CSP plant



Figure 11: Khi Solar One (KSO) is a solar power tower thermal power plant, located in the Northern Cape; Khi Solar One provides 50 megawatts (MW) and is the first solar power plant in Africa, it covers an area of 140 hectares.

Construction of South Africa's largest renewable energy project to date – the Redstone concentrated solar power (CSP) plant – commenced in August 2021. The R11.6 billion solar farm is being built near Postmasburg in the Northern Cape, close to the Jasper and Lesedi photovoltaic (PV) plants and will have a generating output capacity of 100 MW. While this is less than the current largest solar power plant in South Africa – the De Aar project in the Northern Cape – it comes equipped with a 12-hour thermal storage system. This makes it capable of consistently delivering electricity to almost 200 000 households, double the number powered by the De Aar plant. The solar plant is expected to reduce an estimated 440 metric tons of CO_2 emissions per year.

The advantages of solar energy

There are many advantages to solar energy:

- 1. Stable price levels compared to other energy sources: Many energies, especially those that rely on fossil fuels, are expensive since a price is usually determined by supply and demand.
- 2. Cost per energy unit is zero after one-time investment: Solar energy is free. Apart from the upfront cost for the installation of the solar panels and the storage, energy can be produced with no extra cost.
- 3. Solar energy is a renewable energy source.
- 4. Solar power is available all year long and can be stored: Although solar power cannot be produced every day due to clouds or rainfalls, the supply of energy can be assured by using energy storage. By doing so, the energy that is produced during sunny days can be stored for use on days with no sunshine or at night.
- 5. Solar energy will likely become even cheaper in the future: Even though there has already been plenty of research in the field of solar energy, it is likely that technological progress will continue, and this will likely also lead to a decrease in prices for solar panels and batteries for storage.
- 6. Low maintenance costs: Since solar panels are quite durable on average, the maintenance costs are quite low.
- 7. Solar power can be regarded as quite ecologically friendly: Apart from the initial production of solar panels and of solar energy storage, there is virtually no additional pollution associated with the generation of solar energy.

The disadvantages of solar energy

There are however some disadvantages of solar energy:

- 1. Initial production of solar panels produces significant pollution: Although there are no direct environmental costs once the solar panels are set up, the initial installation and production of the solar panels requires the use of precious resources and contributes to several kinds of pollution.
- 2. High upfront costs of solar panels: The upfront costs of solar panels can be quite high at present, as are the batteries needed to store energy.
- 3. Sun intensity varies over the year: Depending on the region, there might be many days or even weeks with rain or clouds and solar energy panels will not be able to generate sufficient electricity.
- 4. Solar farms need large areas of land: For solar farms, large areas of land must be used, which in turn often leads to the destruction of habitats for a variety of animals and plants.

Hydroelectricity

Hydroelectric energy is a form of energy that harnesses the power of water in motion—such as water flowing over a waterfall—to generate electricity.

Most types of hydroelectric power plants have a reservoir of water, a gate or valve to control how much water flows out of the reservoir, and an outlet or place where the water ends up after flowing downward. Water has

gravitational potential energy just before it spills over the top of a dam or flows down a hill. The gravitational potential energy is converted into kinetic energy as water flows downhill. The water can be used to turn the blades of a turbine to generate electricity, which is distributed to the power plant's customers.

There are three different types of hydroelectric energy plants, the most common being an impoundment facility. In an impoundment facility, a dam is used to control the flow of water stored in a pool or reservoir. When more energy is needed, water is released from the dam. Once water is released, gravity takes over and the water flows downward through a turbine. As the blades of the turbine spin, they power a generator.

Another type of plant is called a pumped-storage facility. This plant collects the energy produced from solar, wind, and nuclear power and stores it for future use. The plant stores energy by pumping water uphill from a pool at a lower elevation to a reservoir located at a higher elevation. When there is high demand for electricity, water located in the higher pool is released. As this water flows back down to the lower reservoir, it turns a turbine to generate more electricity.

Eskom has three types of hydroelectric power stations; conventional reservoir – Gariep and Vanderkloof, run-of-river – Colley Wobbles, First and Second Falls and Ncora, and pumped storage schemes – Palmiet, Drakensberg and Ingula.



Figure 12: Gariep Dam in the Free State is one of the largest hydroelectric power stations in South Africa

The advantages of hydroelectricity

There are various advantages of hydroelectricity:

- 1. It is incredibly cheap: Once a power plant goes live and the construction has been completely paid, hydroelectricity is cheaper than all other renewable sources.
- 2. There is a constant source of power: Hydroelectricity is always being produced because water is always

flowing.

- 3. It is a flexible source of power: Many hydroelectricity plants can adapt to changing electrical needs. All they must do is release more water during high periods of electricity demand and fill up the reservoir being maintained during periods of low electricity use. Even on the hottest or coldest days of the year, almost every hydroelectricity plant can meet the demand because of this.
- 4. They add recreational opportunities: Most hydroelectricity plants are based in a dam system on a moving body of water. This means that the reservoir behind the dam must build up for peak use periods and this creates a lake that local communities can use. Beaches, water sports, camping, swimming, and even fishing all become possible and a potential source of new revenues.
- 5. Hydroelectricity plants are highly scalable: Almost every power plant that generates electricity by water flow can be built with a small initial capacity and then be scalable to fit the ongoing needs of the community.
- 6. Hydroelectricity helps fight climate changes and improves air quality: The hydroelectric life cycle produces very small amounts of greenhouse gases. Although only 33% of the available hydroelectric potential has been developed, today hydroelectricity prevents the emission of greenhouse gases corresponding to the burning of 4.4 million barrels of petroleum per day worldwide. Hydroelectric power plants do not release pollutants into the air.

The disadvantages of hydroelectricity

There remain some disadvantages of hydroelectricity:

- 1. Drought conditions can greatly affect output or eliminate it: If the water dries up, then so does the ability to generate power.
- 2. It alters the natural movements of nature: Although modern technology allows for hydroelectric plants to accommodate wildlife movement as much as possible, using salmon runs or alternative migration paths, it is not a perfect system. The filling of a reservoir destroys the habitats of certain animal species, no matter how careful the construction efforts might be.
- 3. They often have a high initial cost: Although the ongoing maintenance costs of hydroelectric power are quite minimal, the initial costs of building a plant require a high level of capital.
- 4. Animals are not the only creatures that need to relocate: Once a dam is installed and the reservoir begins to fill up, there may be people that need to be relocated to new areas as well.
- 5. There are limited places for reservoir development: Many nations have already begun to use up the prime locations for reservoir development that can sustain an ongoing hydroelectric plant.
- 6. Impacts further downstream: The formation of a dam can also have severe impacts on habitats downstream. When the natural flow of a river is blocked so is that of the sediments and nutrients contained within the water. These sediments and nutrients are often vital for replenishing downstream ecosystems such as wetlands, flood plains and deltas. This can lead to the disappearance of birds and other species.

The stagnant reservoir of a hydroelectric dam is often much colder and lower in dissolved oxygen levels. As this water is released through a hydroelectric dam it can alter the temperature and oxygen levels of the water further downstream resulting in dangerous conditions for fish and other river inhabitants.

Sections of river downstream from a hydroelectric dam have the potential to dry out if the flow of water is not managed efficiently. In some countries, hydroelectric power companies are obliged to release set levels of water over the course of the year to mitigate the problem. This careful management can have positive results and reduce the negative impact for habitats and communities further downstream.

As water is held back by a dam, the water table can gradually reduce further downstream. This can have disastrous consequences for plants, shrubs, and trees as their roots struggle to absorb enough nutrients from the soil.

Geothermal power

Geothermal energy is the heat produced deep in the Earth's core. Geothermal energy is a clean, renewable resource that can be harnessed for use as heat and electricity.

Below the Earth's crust, is the core. A small portion of the core's heat comes from the friction and gravitational pull formed when Earth was created more than 4 billion years ago. However, most of the Earth's heat is constantly generated by the decay of radioactive isotopes, such as potassium-40 and thorium-232.

As potassium-40 decays, its nucleus changes, emitting enormous amounts of energy. Radioactive decay is a continual process in the core. Temperatures there rise to more than 5 000 $^{\circ}$ C. Heat from the core is constantly radiating outward and warming rocks, water, gas, and other geological material.

If underground rock formations are heated to about $700 - 1\ 300$ °C, they can become magma. Magma is molten rock permeated by gas and gas bubbles. Magma exists in the mantle and lower crust, and sometimes bubbles to the surface as lava. Magma heats nearby rocks and underground aquifers. Hot water can be released through geysers, hot springs, steam vents, underwater hydrothermal vents, and mud pots.

These are all sources of geothermal energy. Their heat can be captured and used directly for heat, or their steam can be used to generate electricity.

Most of the Earth's geothermal energy does not bubble out as magma, water, or steam. It remains in the mantle, emanating outward at a slow pace and collecting as pockets of high heat. This dry geothermal heat can be accessed by drilling and enhanced with injected water to create steam.



Many countries have developed methods of tapping into geothermal energy. Different types of geothermal energy are available in different parts of the world. In Iceland, abundant sources of hot, easily accessible underground water make it possible for most people to rely on geothermal sources as a safe, dependable, and inexpensive source of energy.

Almost anywhere in the world, geothermal heat can be accessed and used immediately as a source of heat. This heat energy is called low-temperature geothermal energy. Low-temperature geothermal energy is obtained from pockets of heat at about 150 °C. Most pockets of low-temperature geothermal energy are found just a few metres below ground. Low-temperature geothermal energy can be used for heating greenhouses, homes, fisheries, and industrial processes.

Geothermal heat pumps (GHPs) take advantage of the Earth's heat and can be used almost anywhere in the world. GHPs are drilled about 3 to 90 metres deep, much shallower than most oil and natural gas wells. GHPs do not require fracturing bedrock to reach their energy source.

A pipe connected to a GHP is arranged in a continuous loop that circles underground and above ground,

usually throughout a building. The loop can also be contained entirely underground, to heat a parking lot or landscaped area.



1.A refrigerant moves through a loop of pipes.

2. When the weather is cold, the water or refrigerant heats up as it travels through the part of the loop that's buried underground.

3.Once it gets back above ground, the warmed water or refrigerant transfers heat into the building.

4. The refrigerant cools down after its heat is transferred. It is pumped back underground where it heats up once more, starting the process again.

5.On a hot day, the system can run in reverse. The water or refrigerant cools the building and then is pumped underground where extra heat is transferred to the ground around the pipes.

Figure 14: A geothermal heat pump can be installed at individual houses or buildings



1.Hot water is pumped from deep underground through a well under high pressure.

2.When the water reaches the surface, the pressure is

dropped, which causes the water to turn into steam.

3. The steam spins a turbine, which is connected to a generator that produces electricity.

4. The steam cools off in a cooling tower and condenses back to water.

5. The cooled water is pumped back into the Earth to begin the process again.

Figure 15: A flash steam power plant

Flash-steam power plants use naturally occurring sources of underground hot water and steam. Water that is hotter than $_{182}$ °C is pumped into a low-pressure area. Some of the water 'flashes' or evaporates rapidly into steam, and is funnelled out to power a turbine and generate electricity. Any remaining water can be flashed in a separate tank to extract more energy.

Flash-steam power plants are the most common type of geothermal power plants. The volcanically active island nation of Iceland supplies nearly all its electrical needs through a series of flash-steam geothermal power plants. The steam and excess warm water produced by the flash-steam process heat icy sidewalks and parking lots in the frigid Arctic winter.

Geothermal energy is a renewable resource. The Earth has been emitting heat for about 4.5 billion years and will continue to emit heat for billions of years into the future because of the ongoing radioactive decay in the Earth's core. However, most wells that extract the heat will eventually cool, especially if heat is extracted more quickly than it is given time to replenish. Geothermal systems do not require enormous amounts of freshwater. Because water is only used as a heating agent and is not exposed or evaporated. it can be recycled, used for other purposes, or released into the atmosphere as non-toxic steam.

Geothermal energy exists in different forms all over the Earth (by steam vents, lava, geysers, or simply dry heat), and there are different possibilities for extracting and using this heat.
South Africa does not have any geothermal power plants, but according to a study published in 2016, there is huge potential to develop geothermal power plants.

The advantages of geothermal energy

There are several advantages of using geothermal energy:

- 1. Geothermal energy is renewable; it is not a fossil fuel that will be eventually used up. The Earth is continuously radiating heat out from its core and will continue to do so for billions of years.
- 2. Some form of geothermal energy can be accessed and harvested anywhere in the world. Geothermal power plants can last for decades and possibly centuries. If a reservoir is managed properly, the amount of extracted energy can be balanced with the rock's rate of renewing its heat.
- 3. Using geothermal energy is relatively clean. Most systems only emit water vapour, although some emit very small amounts of sulfur dioxide, nitrous oxides, and particulates.
- 4. Unlike other renewable energy sources, geothermal systems can work in the summer or winter, and are not dependent on changing factors such as the presence of wind or sun.
- 5. The space it takes to build a geothermal facility is much more compact than other power plants. To produce a GWh, a geothermal plant uses the equivalent of about 1 046 square kilometres of land. To produce the same GWh, wind energy requires 3 458 square kilometres, a solar photovoltaic centre requires 8 384 square kilometres and coal plants use about 9 433 square kilometres.

The disadvantages of geothermal energy

There are also some disadvantages of using geothermal energy:

- 1. The process of injecting high-pressure streams of water into the Earth can result in minor seismic activity, or small earthquakes.
- 2. Geothermal plants have been linked to subsidence, or the slow sinking of land. This happens as the underground fractures collapse upon themselves. This can lead to damaged pipelines, roadways, buildings, and natural drainage systems.
- 3. Geothermal plants can release small amounts of greenhouse gases such as hydrogen sulfide and carbon dioxide. Water that flows through underground reservoirs can pick up trace amounts of toxic elements such as arsenic, mercury, and selenium. These harmful substances can be leaked into water sources if the geothermal system is not properly insulated.
- 4. Although the process requires almost no fuel to run, the initial cost of installing geothermal technology is expensive.

Biomass

Biomass energy is energy generated or produced by living or once-living organisms. The most common biomass materials used for energy are plants, such as corn and soy. This material can be burned to create heat or converted into electricity.



Figure 16: The most common biomass materials used for energy are plants, wood, and waste; these are called biomass feedstocks

Biomass energy can also be a non-renewable energy source. Biomass contains energy first derived from the sun: plants absorb the sun's energy through photosynthesis and convert carbon dioxide and water into nutrients (carbohydrates). The energy from these organisms can be transformed into usable energy through direct and indirect means. Biomass can be burned to create heat (direct), converted into electricity (direct), or processed into biofuel (indirect).

Thermal conversion

Biomass can be burned by thermal conversion and used for energy. This involves heating the biomass feedstock to burn, dehydrate, or stabilise it. The most familiar biomass feedstocks for thermal conversion are raw materials such as municipal solid waste (MSW) and scraps from paper or lumber mills.

Before biomass can be burned, however, it must be dried. The biomass dries out so completely that it loses the ability to absorb moisture, or rot. It loses about 20% of its original mass but keeps 90% of its energy. During this process, biomass becomes a dry, blackened material. It is then compressed into briquettes. The briquettes have high energy density and are easy to burn during direct or co-firing. Most briquettes are burned directly. The steam produced during the firing process powers a turbine, which turns a generator and produces electricity. This electricity can be used for manufacturing or to heat buildings.

Biomass can also be co-fired or burned with a fossil fuel. Biomass is most often co-fired in coal plants. Cofiring eliminates the need for new factories for processing biomass. Co-firing also eases the demand for coal. This reduces the amount of carbon dioxide and other greenhouse gases released by burning fossil fuels.

Pyrolysis

During pyrolysis, biomass is heated to 300° C without oxygen. This keeps it from combusting and causes the biomass to be chemically altered and produces dark liquid called pyrolysis oil, a synthetic gas called syngas, and a solid residue called biochar. All these components can be used for energy.

Pyrolysis oil, sometimes called bio-oil or biocrude, is a type of tar. It can be combusted to generate electricity and is also used as a component in other fuels and plastics. Syngas can be converted into fuel. It can also be converted into methane and used as a replacement for natural gas.

Biochar is a type of charcoal. Biochar is a carbon-rich solid that is particularly useful in agriculture. Biochar enriches soil and prevents it from leaching pesticides and other nutrients into runoff.

Gasification

Biomass can also be directly converted to energy through gasification. During the gasification process, a biomass feedstock (usually municipal solid waste (MSW)) is heated to more than 700°C with a controlled amount of oxygen. The molecules break down and produce syngas and slag. Syngas is a combination of hydrogen and carbon monoxide. During gasification, syngas is cleaned of sulfur, particulates, mercury, and other pollutants. The clean syngas can be combusted for heat or electricity, or processed into transportation biofuels, chemicals, and fertilisers.

Liquid biofuels

Biomass is the only renewable energy source that can be converted into liquid biofuels such as ethanol and biodiesel. Biofuel is used to power vehicles.

Ethanol is made by fermenting biomass that is high in carbohydrates, such as sugar cane, wheat, or corn. Biodiesel is made from combining ethanol with animal fat, recycled cooking fat, or vegetable oil. Biofuels do not operate as efficiently as gasoline. However, they can be blended with gasoline to efficiently power vehicles and machinery, and do not release the emissions associated with fossil fuels.

Ethanol requires acres of farmland to grow bio crops (usually corn). Growing enough corn for ethanol also creates a strain on the environment because of the lack of variation in planting, and the high use of pesticides.

Hydrogen

Biomass is rich in hydrogen, which can be chemically extracted and used to generate power and to fuel vehicles. Stationary fuel cells are used to generate electricity in remote locations, such as spacecraft and wilderness areas.

Hydrogen fuel cells may hold even more potential as an alternative energy source for vehicles. The U.S. Department of Energy estimates that biomass has the potential to produce 40 million tons of hydrogen per year. This would be enough to fuel 150 million vehicles. Currently, hydrogen fuel cells are used to power buses, forklifts, boats, and submarines, and are being tested on aeroplanes and other vehicles.



Biomass materials such as plants and forests must be sustainably farmed. A steady and varied supply of trees, crops, and other plants is vital for maintaining a healthy environment.



Figure 17: Algae are unique organisms that have enormous potential as a source of biomass energy

Algae

Algae, whose most familiar form is seaweed, produces energy through photosynthesis at a much quicker rate than any other biofuel feedstock. Algae can be grown in ocean water, so it does not deplete freshwater resources. It also does not require soil, and therefore does not reduce arable land that could potentially grow food crops. Although algae releases carbon dioxide when it is burned, it can be farmed and replenished as a living organism. As it is replenished, it releases oxygen, and absorbs pollutants and carbon emissions.

Algae contains oils that can be converted to a biofuel. This creates a 'green crude', which has similar properties to crude oil, and can be used as a biofuel.

The advantages of biomass

There are some important advantages of using biomass:

- 1. Biomass is a clean, renewable energy source. Its initial energy comes from the sun, and plants or algae biomass can regrow in a relatively short amount of time. Trees, crops, and municipal solid waste are consistently available and can be managed sustainably.
- 2. If trees and crops are sustainably farmed, they can offset carbon emissions when they absorb carbon dioxide through respiration. In some bioenergy processes, the amount of carbon that is re-absorbed even exceeds the carbon emissions that are released during fuel processing or usage. Many biomass feedstocks, such as switchgrass, can be harvested on marginal lands or pastures, where they do not compete with food crops.
- 3. Unlike other renewable energy sources, such as wind or solar, biomass energy is stored within the organism, and can be harvested when it is needed.

The disadvantages of biomass

There are several advantages of using biomass:

1. If biomass feedstocks are not replenished as quickly as they are used, they can become non-

renewable. A forest, for instance, can take hundreds of years to re-establish itself.

- 2. Most biomass requires arable land to develop. This means that land used for biofuel crops such as corn and soybeans are unavailable to grow food or provide natural habitats. Biomass fuels also require watering to grow so they are not an ideal source of energy in water scarce countries.
- 3. Forested areas that have matured for decades are able to sequester more carbon than newly planted areas. Therefore, if forested areas are not sustainably cut, re-planted, and given time to grow and sequester carbon, the advantages of using the wood for fuel are not offset by the trees' regrowth.
- 4. Biomass has a lower 'energy density' than fossil fuels. As much as 50% of biomass is water, which is lost in the energy conversion process. Scientists and engineers estimate that it is not economically efficient to transport biomass more than 160 kilometres from where it is processed. However, converting biomass into pellets as opposed to wood chips or larger briquettes can increase the fuel's energy density and make it more advantageous to ship.
- 5. Burning biomass releases carbon monoxide, carbon dioxide, nitrogen oxides, and other pollutants and particulates. If these pollutants are not captured and recycled, burning biomass can create smog and even exceed the number of pollutants released by fossil fuels.

Note

To consolidate your knowledge of renewable energy, you can watch this video 7 Types of Renewable Energy by EcoMastery Project.

7 Types of Renewable Energy (Duration: 08:48)



Table 1: A summary of renewable an	d non-renewable sources of energy
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Energy resource	Energy store	Renewable or non-renewable	Uses	Power output	Impact on environment
Fossil fuels (oil, coal, and natural gases)	Chemical	Non-renewable	Transport, heating, electricity generation	High	Releases CO ₂ (causes global warming)
Nuclear fuels	Nuclear	Non-renewable	Electricity generation	Very high	Radioactive waste (needs to be disposed of safely)
Biofuel	Biofuel	Renewable	Transport, heating, electricity generation	Medium	'Carbon neutral' – little or no effect on the environment. Although growing biofuels can take up land that could be used for farming.
Wind	Kinetic	Renewable	Electricity generation	Very low	Takes up large areas that could be used for farming; windmills change the landscape
Hydroelectricity	Gravitational potential	Renewable	Electricity generation	Medium	Local habitats are affected by the large areas that need to be flooded to build dams
Geothermal	Internal (thermal)	Renewable	Electricity generation, heating	Medium	Very low
Tides	Kinetic	Renewable	Electricity generation	Potentially very high, but hard to harness	Tidal barrages can block sewage which needs to go out to sea; local habitats also affected
Sun	Nuclear	Renewable	Electricity generation, heating	Dependent on the weather and only available during daylight	Very little
Water waves	Kinetic	Renewable	Electricity generation	Low	Very low



Exercise 1.2

- 1. Which of the following groups are all renewable energy sources?
 - a. The tides, water waves and nuclear fuel
 - b. Biofuel, geothermal and coal
 - c. Geothermal, the tides and hydroelectricity
- 2. A learner suggests replacing a coal power station with solar panels on the same site. Give two reasons why this might not be good idea.
- 3. Give three ways of generating electricity in which no fuel is burned, and the energy is renewable.
- 4. Petrol and ethanol are both fuels. Petrol is made from oil. Oil could run out in 100 years.

In some countries people plant sugar cane and use it to make ethanol, which can be used as a fuel for transportation.

- a. What type of renewable fuel is sugar cane?
- b. Sugar cane will not run out. Explain why.

The <u>full solutions</u> are at the end of the unit.

Electrochemical cells

A fuel cell can be defined as an electrochemical cell that generates electrical energy from fuel via an electrochemical reaction. These cells require a continuous input of fuel and an oxidising agent (generally oxygen) to sustain the reactions that generate the electricity. Therefore, these cells can constantly generate electricity until the supply of fuel and oxygen is cut off.

Despite being invented in the year 1838, fuel cells began commercial use when NASA used them to power space capsules and satellites. Today, these devices are used as the primary or secondary source of power for many facilities including industries, commercial buildings, and residential buildings.



Figure 18: A fuel cell consists of a cathode, an anode, and an electrolyte; in these cells, the electrolyte enables the movement of the protons

The reaction between hydrogen and oxygen can be used to generate electricity via a fuel cell. Such a cell was used in the Apollo space programme, and it served two different purposes – It was used as a fuel source as well as a source of drinking water (the water vapour produced from the cell, when condensed, was fit for human consumption).

The working of this fuel cell involved the passing of hydrogen and oxygen into a concentrated solution of sodium hydroxide via carbon electrodes. The cell reaction can be written as follows:

Cathode reaction: $O_2\,+\,2H_2O\,+\,4e^{-}\rightarrow4OH^{-}$

Anode reaction: $2H_2 + 4OH^-
ightarrow 4H_2O + 4e^-$

Net cell reaction: $2H_2\,+\,O_2~\rightarrow 2H_2O$

However, the reaction rate of this electrochemical reaction is quite low. To overcome this issue, a catalyst such as platinum or palladium is used. To increase the effective surface area, the catalyst is finely divided before being incorporated into the electrodes.

The efficiency of the fuel cell described above in the generation of electricity generally approximates to 70% compared to the efficiency of thermal power plants at 40%. This difference is because the generation of electric current in a power plant involves the conversion of water into steam, and the usage of this steam to rotate a turbine. Fuel cells, however, directly convert chemical energy into electrical energy.

The Department of Science and Technology, wants South Africa to achieve 25% of the global market in the production of hydrogen and fuel cell catalysts, using platinum group metal catalysts. With platinum being a key component for most fuel cells, the country is well placed to do so, particularly as it has three quarters of the world's platinum reserves.

Types of fuel cells

There are several varieties of fuel cells:

The Polymer Electrolyte Membrane (PEM) Fuel Cell used for transport and stationary and portable generation of power.

The Solid Acid Fuel Cell which uses solid acids including caesium hydrogen sulfate and caesium dihydrogen phosphate respectively.

The Alkaline Fuel Cell which uses an aqueous alkaline solution. These cells are highly efficient. They also produce heat and water along with electricity.

The Solid Oxide Fuel Cell uses a solid oxide or a ceramic electrolyte. These fuel cells are highly efficient and have a relatively low cost.

The Molten Carbonate Fuel Cell uses lithium potassium carbonate salt. These cells can be powered by carbon-based fuels such as natural gas and biogas.

Applications of fuel cells

There are a number of applications of fuel cells:

- 1. Power: Stationary fuel cells are used for commercial, industrial, and residential primary and backup power generation. Fuel cells are very useful as power sources in remote locations, such as spacecraft, remote weather stations, large parks, communications centres, rural locations, and in certain military applications. A fuel cell system running on hydrogen can be compact and lightweight and have no major moving parts. Because fuel cells have no moving parts and do not involve combustion, in ideal conditions they can achieve up to 99.9999% reliability.
- 2. Co-generation: Combined heat and power (CHP) fuel cell systems, including micro combined heat, power office building and factories. The system generates constant electric power (selling excess power back to the grid when it is not consumed), and at the same time produces hot air and water from the waste heat. The waste heat from fuel cells can be diverted during summer directly into the ground providing further cooling while the waste heat during winter can be pumped directly into the building.
- 3. Fuel cell electric vehicles (FCEVs): A fuel cell vehicle (FCV) or fuel cell electric vehicle (FCEV) is an electric vehicle that uses a fuel cell, sometimes in combination with a small battery or supercapacitor, to power its onboard electric motor. Fuel cells in vehicles generate electricity generally using oxygen from the air and compressed hydrogen. Most fuel cell vehicles are classified as zero-emission vehicles that emit only water and heat.



Figure 19: Fuel cell electric vehicles, or FCEVs, use clean fuels and are therefore more eco-friendly than internal combustion engine-based vehicles

This technology is also being used to power buses, forklifts, motorcycles, boats, and aeroplanes.

These vehicles differ from electric vehicles, which use rechargeable batteries.

4. Portable power systems: As power demands for cell phones increase, fuel cells could become much more attractive options for larger power generation. The demand for a longer battery life for mobile phones and computers is something often demanded by consumers so fuel cells could start to make strides into laptop and cell phone markets.

Fuel cells for use in larger scale operations also show much promise. Portable power systems that use fuel cells can be used in the leisure sector, the industrial sector (i.e. power for remote locations including gas/oil well sites, communication towers, security, weather stations), and in the military sector. The key advantage of fuel cells in this market is the great power generation per weight. While fuel cells can be expensive, for remote locations that require dependable energy, fuel cells provide efficient power solutions.

The advantages of fuel cells

Fuel cells have several advantages:

- 1. Renewable and readily available: Hydrogen is the most abundant element in the Universe and despite the challenges associated with its extraction from water, it is a uniquely abundant and renewable source of energy.
- 2. Hydrogen is a clean and flexible energy source to support zero-carbon energy strategies: Hydrogen fuel cells provide an inherently clean source of energy, with no adverse environmental impact during operation as the by-products are simply heat and water (which you can drink).

Hydrogen fuel cells do not generate greenhouse gas emissions like fossil fuel sources, thus reducing pollution and improving air quality as a result.

- 3. More powerful and energy efficient than fossil fuels: Hydrogen fuel cell technology provides a highdensity source of energy with good energy efficiency. Hydrogen fuel cells are more efficient than many other energy sources, including many green energy solutions.
- 4. Fast charging times: The charge time for hydrogen fuel cell power units is like that for conventional internal combustion engine vehicles and much quicker in comparison to electric vehicles which can take several hours to charge, hydrogen fuel cells can be recharged in under five minutes.
- 5. No noise pollution
- 6. Long usage times: A hydrogen vehicle has the same range as those that use fossil fuels.
- 7. Ideal for use in remote areas: The availability of hydrogen through local generation and storage could prove to be an alternative to diesel-based power and heating in remote areas.
- 8. Democratisation of power supply: Hydrogen fuel cells have the potential to reduce the dependency of a nation on fossil fuels, which will help democratise energy and power supplies around the world. This increased independence will prove a benefit for many countries who are currently reliant on fossil fuel supply.

The disadvantages of fuel cells

There are some notable disadvantages to using fuel cells:

- 1. Hydrogen extraction: Despite being the most abundant element in the Universe, hydrogen does not exist on its own so needs to be extracted from water via electrolysis or separated from carbon fossil fuels. Both processes require a significant amount of energy to achieve. In addition, this extraction typically requires the use of fossil fuels, which reduces the green credentials of hydrogen.
- 2. Investment is required: Hydrogen fuel cells need investment to be developed to the point where they become a genuinely viable energy source.
- 3. Cost of raw materials: Precious metals such as platinum and iridium are typically required as catalysts in fuel cells and some types of water electrolyser, which means that the initial cost of fuel cells can be high. This high cost has deterred some from investing in hydrogen fuel cell technology.
- 4. Overall cost: The cost for a unit of power from hydrogen fuel cells is currently greater than other energy sources, including solar panels.
- Hydrogen storage and transportation: Hydrogen is a highly flammable fuel source, which brings understandable safety concerns. Hydrogen gas burns in air at concentrations ranging from 4% to 75%. Storage and transportation of hydrogen is more complex than that required for fossil fuels. This implies additional costs to consider for hydrogen fuel cells as a source of energy.
- 6. Infrastructure: Because fossil fuels have been used for decades, the infrastructure for this power supply already exists. Large scale adoption of hydrogen fuel cell technology for vehicles will require new refuelling infrastructure to support it.

Note

To further understand fuel cells you can watch this video called Hydrogen & Fuel Cells by Fuse Schools.

Hydrogen & Fuel Cells (Duration: 5.00)





Exercise 1.3

- 1. True or false: A fuel cell is a type of electrochemical cell?
- 2. Which of the following can be used as fuel in a fuel cell?
 - a. Nitrogen
 - b. Helium
 - c. Carbon
 - d. Hydrogen
- 3. Which of the following is not produced by a fuel cell?
 - a. Electricity
 - b. Greenhouse gases
 - c. Heat
 - d. Water
- 4. Which of the following statements are false:
 - a. Hydrogen is highly flammable.
 - b. Fuel cells do not need refuelling and will not run out of generating capacity.
 - c. The cost of the catalysts used is very high.
 - d. Fuel cells do not cause pollution.

The <u>full solutions</u> are at the end of the unit.

Summary

In this unit you have learnt the following:

- Sources of energy can be renewable or non-renewable.
- Non-renewable sources of energy are fossil fuels (coal, oil and gas) and nuclear.
- · Coal, oil and gas are burnt to release energy to generate electricity and power vehicles.
- Burning fossil fuels releases greenhouse gases which contribute to climate change and gases which contribute to acid rain.
- Nuclear energy uses uranium as a fuel.
- Nuclear energy is very efficient and does not release greenhouse gases or gases which contribute to acid rain.
- Renewable sources of energy will not run out.

- Renewable sources of energy include wind and solar.
- Renewable sources of energy do not release greenhouse gases or gases which contribute to acid rain.

Unit 1: Assessment

Suggested time to complete: 30 minutes

- 1. Which of these is a non-renewable energy resource?
 - a. Coal
 - b. Geothermal
 - c. Wind
- 2. What does wind directly turn to generate electricity?
 - a. Generator
 - b. Motor
 - c. Turbine
- 3. Which of these is an advantage of wind energy?
 - a. There is no electricity when there is no wind
 - b. No harmful gases are produced
 - c. Turbines may be noisy
- 4. Which of these is a disadvantage of hydroelectric power?
 - a. It is very reliable
 - b. Dams destroy the habitats of estuary species
 - c. There are no fuel costs
- 5. Which type of energy does not come from the Sun?
 - a. Biomass
 - b. Wind
 - c. Geothermal
- 6. Which of these is a disadvantage of geothermal energy?
 - a. Power stations can only be built in certain areas
 - b. It is renewable
 - c. No harmful gases are produced
- 7. What do photovoltaic panels generate?
 - a. Electricity
 - b. Steam
 - c. Hot water
- 8. Which of these is a disadvantage of photovoltaic panels?
 - a. They are renewable
 - b. They do not work at night
 - c. They only heat water slowly
- 9. Where does the energy come from in geothermal power stations?
 - a. Fuel

- b. The Sun
- c. The Earth's hot core
- 10. The national grid is supplied with electricity from large-scale electrical generators. These generators may be driven using different energy sources. Compare the use of a non-renewable energy source with the use of a renewable energy source to produce electricity for the national grid.
- 11. Wind energy requires significant amounts of empty land where wind blows regularly. Suppose a proposal is made to turn most of the Western and Eastern Cape into a gigantic wind farm. Would you support the proposal? Please justify your answer.
- 12. Imagine you were the Minister for Environmental Affairs, and the President has asked you to reduce the amount of fossil fuels used in South Africa. Which methods would you use to generate electricity in South Africa? Justify your answer.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

- 1. Advantage: energy efficient, cleaner than fossil fuels, low running costs, fuel costs are low. Disadvantages: disposal/storage of radioactive waste, potential for radioactive disaster!
- 2. Non-renewable sources of energy, burning them emits pollutants which contribute to global warming.

3.

- abundant supplies of coal
- infrastructure to generate electricity
- cheap fuel.

Back to Exercise 1.1

Exercise 1.2

1. c

- 2. Solar panels will not provide a consistent supply of energy as the coal power station does. The amount of energy supplied by the solar panels per hour will be much lower than the amount supplied by the coal power station.
- 3. any 3 of the following:
 - solar
 - wind
 - geothermal
 - hydroelectric
 - wave
 - tidal

4.

a. biomass

b. Sugar cane can be replanted again and again.

Back to Exercise 1.2

Exercise 1.3

- 1. True
- 2. d
- 3. b
- 4. b

Back to Exercise 1.3

Unit 1: Assessment

- 1. a
- 2. c
- 3. b
- 4. b
- 5. c
- 6. a
- 7. a
- 8. b
- 9. c
- 10. First, you should write down the two energy sources you have chosen, and say which is renewable and which is non-renewable, e.g. 'Wind is a renewable energy source. Coal is a non-renewable energy source'. Then comment on at least two similarities and two differences between the energy sources.

Fossil fuel power stations are cheaper to build than wind farms for the same power output. However, fossil fuel power stations produce CO_2 which may increase global warming – renewable energy generators (wind farms) do not.

Renewable energy generators have a free or cheaper source of fuel.

Fossil fuels must be taken out of the ground.

Nuclear power stations produce radioactive waste, which is dangerous, none of the other energy generators do this.

Wind, waves, and sun are unreliable sources of energy, but fossil and nuclear fuels are always available.

11. Either 'yes you would approve the proposal' or 'no you wouldn't' needs to be stated and then reason given for your choice.

No:

- Affects bird populations.
- Wind is not reliable.
- Wind turbines cause visual pollution and noise pollution there is a large tourist industry in the Western Cape which would be affected by the sight and noise of the turbines.
- Wind farms may destroy landscapes and habitats, as trees need to be removed to build the

turbines.

• Wind turbines are expensive.

Yes:

- Renewable source of energy.
- Does not release polluting gases, so does not contribute to climate change.
- Farmers can still farm below the turbines.
- Fuel costs are zero because the wind is free.
- Low maintenance costs of the turbines.
- 12. Any renewable source is fine, but remember to justify your answer:

Solar: The abundance of sunshine in South Africa (makes this the best choice). Large parts of the Northern Cape and the Karoo are uninhabited, so there is lots of space. Cities and towns could have solar panels installed on roof tops, and in parking lots to allow for solar farms. It is a clean, and renewable source of energy. Once the farms have been set up, the fuel is free. Little maintenance costs, no pollutants and in areas such as the Northern Cape and the Karoo there are very few days in a year when it is cloudy, so it is a reliable source of energy.

Wind: Renewable source of energy Does not release polluting gases, so does not contribute to climate change. Farmers can still farm below the turbines. Fuel costs are zero because the wind is free. Low maintenance costs of the turbines.

Nuclear: South Africa has an abundant supply of Uranium although it is a non-renewable source of energy. Maintenance costs once the power stations have been built are low. It is an efficient and reliable source of energy. Does not need as much space as a solar or wind farm and is a cleaner fuel than using fossil fuels.

Biomass: not really an option for SA because this is a water scarce country. Geothermal: can be used but is extremely expensive. Hydroelectricity: can be used, but there are cheaper, more efficient, more environmentally friendly options available.

Back to Unit 1: Assessment

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SUBJECT OUTCOME XV CHEMICAL SYSTEMS AND INDUSTRY: IDENTIFY AND CRITICALLY EVALUATE THE IMPACT OF SCIENTIFIC KNOWLEDGE ON THE CHEMICAL INDUSTRIES AND THE QUALITY OF HUMAN, ENVIRONMENTAL AND SOCIO-ECONOMIC

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Subject outcome

Subject outcome 7.2: Identify and critically evaluate the impact of scientific knowledge on the chemical industries and the quality of human, environmental and socio-economic



Learning outcomes

- Study the following industrial examples with appropriate topics in terms of resources, needs and the chemical connection:
 - Organics: Sasol, Polyfin (monomer and polymers); petroleum industry, paint, and adhesive industry.
 - Acids and alkali: Chloralkali (soap, PVC, etc), cement (Lafarge and PPC), mines (extraction of ore), explosives (AECI), fertiliser industry (N, P, K).
 - Redox reactions: Electrochemical and battery industries. Identify and critically evaluate the impact of scientific knowledge on the chemical industries and the quality of human, environmental and socio-economic development.



Unit 1 outcomes

By the end of this unit you will be able to evaluate and identify the impact of scientific knowledge on certain industries in South Africa including:

- Organic chemistry: petroleum industry, paint, and adhesive manufacture.
- The production of soaps, cement, mining, the manufacture of explosives and fertiliser.
- Electrochemical and battery industries.

Unit 1: Impact of Scientific Knowledge on Industry

EMMA HARRAGE



What you should know

Before you start this unit, make sure you can:

- Understand organic chemistry. Refer to level 4 subject outcome 5.2 units 1 if you need help with this.
- Understand redox reactions. Refer to level 4 subject outcome 6.2 unit 1 if you need help with this.
- Understand acids and alkalis. Refer to level 4 subject outcome 6.1 unit 1 if you need help with this.
- Understand chemical equilibrium. Refer to level 4 subject outcome 6.3 unit 1 if you need help with this.

Introduction

South Africa has developed an established, diversified manufacturing base which provides employment and economic empowerment.

Manufacturing in South Africa

Manufacturing in South Africa is dominated by the following industries:

Agri processing: Agriculture contributes 4% to South Africa's gross domestic product (GDP) and consists of cattle and sheep farming, with only 13% of land used for growing crops. Maize is most widely grown followed by wheat, oats, sugar cane and sunflowers. Citrus and deciduous fruits are exported, as are locally produced wine and flowers.

Automotive: The automotive industry is one of South Africa's most important sectors, with many of the major multinationals using South Africa to source components and assemble vehicles for both the local and international markets.

Chemicals: The synthetic coal and natural gas-based liquid fuels and petrochemicals industry is prominent, with South Africa being a world leader in coal-based synthesis and gas-to-liquids (GTL) technologies. South Africa's chemical industry is of substantial economic significance to the country, contributing around 5% to

the gross domestic product (GDP) and approximately 25% of its manufacturing sales. Sasol (through Sasol Chemical Industries and Sasol Polymers) is the most dominant, followed by AECI and Dow Sentrachem.

Industry & Mining



Figure 1: The location of the major industries in South Africa

ICT and electronics: South African software developers are recognised as world leaders in innovation, production and cost efficiency backed by an excellent local infrastructure. This sector can be divided into three main sub-sectors: telecommunications, electronics, and information technology.

Metals: South Africa's large, well-developed metals industry, with vast natural resources and a supportive

infrastructure, represents a third of all South Africa's manufacturing. It comprises basic iron ore and steel, basic non-ferrous metals, and metal products. The iron and steel basic industries involve the manufacture of primary iron and steel products from smelting to semi-finished stages.

Textiles, clothing, and footwear: Due to technological developments, local textile production has evolved into a capital-intensive industry, producing synthetic fibres in ever-increasing proportions.

The polyfins industry in South Africa

A polyfin is a type of polymer with the general formula $(CH_2CHR)_n$. They are made from a handful of simple alkenes. Polyethylene and polypropylene are the most used commercially. More specialised polyfins include polybutene, polyisobutylene and polymethyl pentene.

Thermoplastic polyfins include the following:

- low-density polyethylene (LDPE)
- linear low-density polyethylene (LLDPE)
- very-low-density polyethylene (VLDPE)
- ultra-low-density polyethylene (ULDPE)
- medium-density polyethylene (MDPE)
- polypropylene (PP)
- polymethyl pentene (PMP)
- polybutene-1
- ethylene-octene copolymers
- stereo-block PP
- olefin block copolymers
- propylene–butane copolymers.

Polyolefin elastomers (POE) include the following:

- polyisobutylene (PIB),
- poly(a-olefin)s,
- ethylene propylene rubber (EPR),
- ethylene propylene diene monomer (M-class) rubber (EPDM rubber).

Polyfin properties range from liquidlike to rigid solids and are primarily determined by their molecular weight and degree of crystallinity. Polyfin surfaces have excellent chemical resistance and are unaffected by common solvents. They can be adhesively bonded and by some superglues. They are extremely inert chemically but show less strength at lower and higher temperatures.

The major applications for HDPE are film (wrapping of goods), blow moulding (e.g. liquid containers, such as bleach bottles), injection moulding (e.g., toys, screw caps), extrusion coating (e.g. coating on milk cartons), piping for distributing water and gas, insulation for telephone cables, wire and cable insulation.

LLPE is mainly (70%) used for film.

Major applications of polypropylene are injection moulding, fibres, and film. Compared to polyethylene, polypropylene is stiffer but less prone to breaking. It is less dense and more chemically resilient.

Petrochemical industry

South Africa's fuel sector, including petrol, diesel, jet fuel, illuminating paraffin, fuel oil, bitumen, and liquefied petroleum gas, contributes about 8.5% to the country's GDP while supplying about 18% of its primary energy. South Africa has six refineries, although three of them are currently shut and in need of investment. South Africa has a trade deficit in petrol and diesel, driven by demand from Eskom's open cycle

gas turbines, and demand from other sectors. South Africa is expected to become more dependent on fuel imports as the country's six refineries face an uncertain future.

Two of the main companies involved in the petroleum industry are:

PetroSA: The Petroleum Oil and Gas Corporation of South Africa (SOC) Limited (PetroSA) is the national oil company of South Africa which is owned by the State and reports to the Department of Energy.

The core business activities of PetroSA are:

- The exploration and production of oil and natural gas.
- The production of synthetic fuels from offshore gas at one of the world's largest Gas-to-

Liquid (GTL) refineries in Mossel Bay, South Africa.

- The development of domestic refining and liquid fuels logistical infrastructure.
- The marketing and trading of oil and petrochemicals.

PetroSA operates the FA-EM, South Coast gas fields as well as the Oribi and Oryx oil fields. These gas fields provide feedstock to the Mossel Bay GTL refinery. PetroSA's GTL refinery produces synthetic fuels and products converted from natural methane-rich gas and condensate using a unique GTL Fischer-Tröpsch technology. Key commodities produced include unleaded petrol, kerosene (paraffin), diesel, propane, liquid oxygen and nitrogen, distillates, eco-fuels, and alcohols.

SASOL: Sasol Mining operates six coal mines that supply feedstock for Secunda (Sasol Synfuels) and Sasolburg (Sasolburg Operations) complexes in South Africa. While the coal supplied to Sasol Synfuels is used as gasification feedstock, some is used to generate electricity. The coal supplied to the Sasolburg Operations is used to generate electricity and steam.

Sasol operations include fuel blending and storage facilities at the Secunda operations to turn components into market ready products, such as:

- bitumen
- industrial heating fuels
- naphtha, and illuminating paraffin
- transport fuels, (petrol, diesel and jet fuels)
- lubricants and lubricant base oils
- liquefied petroleum gas (LPG)
- automotive lubricants
- industrial lubricants
- greases
- cleansers and degreasers
- \cdot automotive fuels
- burner fuels

Fischer-Tröpsch processes

The Fisher-Tröpsch process is used to create chemicals called alkanes. Sometimes the gas methane is produced, which is generally undesirable. Sometimes, different kinds of alcohols are produced in small amounts.

Sasol's Fischer-Tröpsch hard waxes are synthetically produced by using gas-to-liquids (GTL) technology, in which Sasol is globally recognised as a commercial and technical pioneer. High-Temperature Fischer-Tröpsch (or HTFT) is performed at temperatures of 330°C - 350°C. HTFT uses an iron-based catalyst. Sasol uses HTFT in Coal-to-Liquid plants (CTL).



Figure 2: A flow diagram of the Fisher-Tröpsch process; natural gas is combined with oxygen to form a syngas which is then subjected to a Fischer-Tröpsch process conversion, resulting in waxy synthetic crude; finally, this is cracked down to produce the product

Many different catalysts can be used for the Fischer-Tröpsch process. The most common catalysts are the metals cobalt, iron, and ruthenium. Cobalt is the most active catalyst (it has the greatest and fastest effect on the process). When the input is a natural gas, cobalt catalysts are very good for the Fischer-Tröpsch process. Iron catalysts are better when the input gas is of lower quality (less pure) such as coal or biomass.

Most metals used for this process (such as cobalt, nickel, and ruthenium) remain in their metal form when added to the process. However, iron catalysts behave very differently. Often, iron catalysts change form and chemical phase, like converting into various oxides and carbides during the reaction. It is important to control all of the iron reactions during the process, or else the process may not work correctly.

The GTL process uses natural gas, to produce hard waxes. With the combination of high melting point, low viscosity, and excellent hardness even at high temperatures, hard waxes provide a useful variety of applications. The Sasol process turns natural gas into energy and chemical products, including transport fuels, base oils, waxes, paraffins and naphtha.

The liquid fuels produced are high-performance, low-emission products. The synthetic GTL-based diesel is an environmentally cleaner burning fuel as it leads to a reduction in carbon monoxide, hydrocarbon, and particulate matter without compromising NOx emissions even when compared to European sulfur-free diesel.



Sasol was founded in 1950 by the state-owned Industrial Development Corporation to produce gasoline from hard coal based on the Fischer-Tröpsch synthesis. The technique of coal liquefaction was

considered unprofitable at the time and was used because South Africa was subject to extensive sanctions during the apartheid era, which made it difficult to source crude oil, while domestic coal could be extracted cheaply. Production began in 1955.

The company owns plants in 17 countries on three continents, namely in Belgium, Brazil, the People's Republic of China, Germany, Dubai, France, Italy, Japan, Netherlands, Poland, Russia, Singapore, Slovak Republic, Spain, South Africa, United States, and the United Kingdom.



Sasol's primary business is the further processing of coal and natural gas by means of coal gasification and Fischer-Tröpsch synthesis into gasoline and raw materials for the chemical industry in the Sasol 1 (Sasolburg) and Sasol 2 and 3 plants (both in Secunda). The low-temperature syngas conversion process is operated based on an iron-based catalyst to produce hydrocarbon waxes and paraffins. The syngas (mixture of hydrogen and carbon monoxide) is also converted into methanol, butanol, and ammonia. Ammonia is then converted into nitric acid and ammonium-based fertilisers and explosives.

The process also produces water and oxygenated hydrocarbons ethylene and ethane, which are then purified and marketed. The ethane is further cracked to make additional ethylene which is then converted into polyethylene for the polymer business. Propylene from the Secunda and Sasolburg plants are also converted into products such as polypropylene, butanol, butyl acrylate and ethyl acrylate. The acrylates are used to make superabsorbent polymers which are used in nappies. Through proprietary technology 1-hexene, 1-octene and 1-pentene are recovered from the oil stream. International customers use these as co-monomers for making speciality grade polymers.

The South African energy cluster operates coal mining in South Africa (Sasol mining), trades in natural gas from Mozambique (Sasol gas), operates the CtL plant in Secunda (Sasol synfuels), an oil refinery and processing plant, as well as various sales activities including filling stations (Sasol oil).



Exercise 1.1

- 1. Using <u>figure 2</u>, answer the following questions:
 - a. What is the first step in the process of turning coal into liquid fuels?
 - b. Which catalyst is used in Sasol's Fischer-Tröpsch process? Why is that one used?
- 2. What happens to the heavy waxes produced in the Fischer-Tröpsch process?
 - a. They are separated out
 - b. They are mixed with gasoline
 - c. They are subjected to cracking
 - d. No heavy oil fractions are produced

The <u>full solutions</u> are at the end of the unit.

Paint and adhesives

South Africa has many manufacturers of paint and adhesives, including Plascon, Prominent Paints, Durham paints and Universal. Probably the most well-known adhesive manufacturer in South Africa is Pratley's. Other major manufacturers include Alcolin, and Pattex.

Paint manufacture

The South African paint industry is valued at R11.5 billion per annum. The industry supports a value chain worth R150 billion, employs approximately 10 000 people and is an essential supplier to the construction and manufacturing sectors as many products cannot be sold without a protective coating.

A paint is composed of pigments, solvents, resins, and various additives. The pigments give the paint colour; solvents make it easier to apply; resins help it dry; and additives serve as everything from fillers to anti-fungicidal agents. The basic white pigment is titanium dioxide, selected for its excellent concealing properties, and black pigment is commonly made from carbon black. Other pigments used to make paint include iron oxide and cadmium sulfide for reds, metallic salts for yellows and oranges, and iron blue and chrome yellows for blues and greens.



Solvents are various low viscosity, volatile liquids. They include petroleum mineral spirits and aromatic solvents such as benzol, alcohols, esters, ketones, and acetone. The natural resins most used are linseed, coconut, and soybean oil, while alkyds, acrylics, epoxies, and polyurethanes are among the most popular synthetic resins. Calcium carbonate and aluminium silicate are simply fillers that give the paint body and substance without changing its properties. Other additives produce certain desired characteristics in paint, such as the thixotropic agents that give paint its smooth texture, driers, anti-settling agents, anti-skinning agents, defoamers, and a host of others that enable paint to cover well and last long.

Adhesives

Adhesives, also known as glue, cement, mucilage, or paste, are any non-metallic substance applied to one or both surfaces of two separate items that binds them together and resists their separation.

The use of adhesives offers certain advantages over other binding techniques such as sewing, mechanical fastenings, or welding. These include the ability to bind different materials together, the more efficient distribution of stress across a joint, the cost-effectiveness of an easily mechanised process, and greater flexibility in design. Disadvantages of adhesive use include decreased stability at high temperatures, relative weakness in bonding large objects with a small bonding surface area, and greater difficulty in separating objects during testing.

Adhesives are typically organised by the method of adhesion followed by reactive or non-reactive, a term which refers to whether the adhesive chemically reacts to harden. Adhesives may be found naturally or produced synthetically.

There is hardly any product in our surroundings that does not contain at least one adhesive—be it the label on a beverage bottle, protective coatings on automobiles, or profiles on window frames. Market researchers forecasted a turnover of almost US 50 billion for the global adhesives market in 2019.

Adhesive type	Sources/properties	Common uses	
Animal glue	Obtained from animal by-products such as bones, blood, and hooves	Binding of abrasives in sandpaper and other grinding materials	
Casein	Main protein in milk	Labels on beer bottles that do not come off in ice water, yet are recyclable	
Starch	From corn and maize	Corrugated cardboard bonding	
Natural rubber	Not 'sticky enough' by itself but is used as an additive in other adhesives	Self-adhesive envelopes and other pressure-sensitive adhesives; adhesives that bond to substrates on contact (such as tapes)	
Butyl rubber/isobutylene	lt is elastomeric – It stretches	Additive for hot-melt adhesives, window sealants, and pressure-sensitive adhesives	
Amino resins	Water-soluble adhesives	Bonding of layers in plywood and the bonding of particles in particle board	
Polyurethane	A flexible adhesive	Bonding soles to the bodies of shoes; also used in food packaging	
Polyvinyl acetate	Common 'white' glue	Book bindings and labels	
Polyolefin/ethylene copolymer	No solvents involved	Hot melts: these are thermoplastics applied in molten form (in the range) which solidify on cooling to form strong bonds between a wide range of materials; ethylene-vinyl acetate-based hot-melts are particularly popular for crafts because of their ease of use and the wide range of common materials they can join	
Acrylates or anaerobic adhesives	Cure when air is removed	Adhesive used to keep nuts tight on bolts, such as those within ATMs and heavy machinery	
Silicone	Both an adhesive and a sealant and the only common adhesive that is based on silicon rather than carbon	Bathtub and shower sealants; also many car applications, such as oil pans and head gaskets	

Table 1: Adhesives are typically organised by the method of adhesion

Note

The only South African product which has been to space is Pratley Putty. Pratley Putty is a clay-like adhesive dough designed by South African engineer George Montague Pratley from Krugersdorp. He discovered Pratley Putty in the 1960s while searching for a glue with which he could paste components into an electrical split box. Pratley Putty was used in 1969 to hold together parts of the Eagle moon

landing craft of the Apollo 11 mission. It was the first to land on the moon. It is the only South African product to land on the moon.

Mining in South Africa

Minerals are natural compounds formed through geological processes. A mineral could be a pure element, but more often minerals are made up of many different elements combined. Minerals are useful chemical compounds for making new materials that we can use in our daily lives. Mining is an important industry in South Africa. We have a lot of mineral resources in our country and a lot of people depend on mining for a living.



Leading mining companies in Africa based on number of employees in 2020

Figure 4: The mining industry is an important part of the South African economy, contributing 8.2% of the country's total GDP in 2020; South Africa's mining industry employed a total of 451 427 people in 2020

Long before diamonds were discovered in the Kimberley area and the Gold Rush in the Pilgrim's Rest and Witwatersrand areas in the late 1800s, minerals have been mined in South Africa. At Mapungubwe in the Limpopo Province evidence of gold and iron mining and smelting was found which dates to the early 11th century AD. However, it was the large-scale mining activities that accelerated the development of the country.



Figure 5: The Bushveld Igneous Complex has the world's largest primary source of platinum group metals, indicated on the map in light blue; it is one of the most important mining areas in South Africa due to its abundance of minerals

South Africa has a wealth of minerals. We are the world's largest producers of chromium, manganese, platinum, vanadium and andalusite; and the second largest producer of ilmenite, palladium, rutile, and zirconium. We are the third largest coal exporter, fifth largest diamond producer and seventh largest iron ore producer. Up to 2010 we were the world's largest gold producer, but our gold production has declined steadily over a number of years. We are currently fifth on the list of gold producers.

Leading mining companies in Africa based on market capitalization as of August 2021 (in billion U.S. dollars)



Figure 6: The leading mining companies in South Africa

Mining has played a major role in the history of South Africa. It accelerated technological development and created infrastructure in remote areas in South Africa. Many small towns in South Africa started because of mining activity in the area. It also created a demand for roads and railways to be built. Most importantly it created job opportunities for thousands of people. Even today many households are dependent on the mining activities for jobs and an income. Mining is an important part of our economic wealth. We export minerals and ore to many other countries in the world.

• **Extraction of iron:** South Africa is the seventh largest producer of iron ore in the world. Iron has been mined in South Africa for thousands of years. Evidence of early mining activities was found in archaeological sites dating back to the Iron Age around 770 AD.

Iron is extracted from ore using a blast furnace. Iron ore, a type of coal called coke (which contains 85% carbon) and lime are added to the top of the blast furnace. Hot air provides the oxygen for the reaction. The temperature of a blast furnace can be up to $1\ 200\$ °C. The reaction takes place inside the furnace and molten iron is removed from the bottom. Lime (calcium carbonate) is added to react with the unwanted materials, such as sand (silicon dioxide). This produces a waste product called slag. The slag is removed from the bottom and used for building roads. Iron is used to make steel. In 2020, South Africa's production of iron ore amounted to an estimated 71 million metric tons. South Africa is one of the world's largest producers of iron ore.



Figure 7: A blast furnace

- Gold: One of the largest gold companies in the world, AngloGold Ashanti, headquartered in Johannesburg, was founded in 2004. AngloGold Ashanti generated revenues of 4.43 billion in 2020. It produced over three million ounces of gold in 2020, with over half of it being produced in Africa. Two of the deepest gold mines worldwide are situated in South Africa. These are the AngloGold Ashanti's Mponeng gold mine that is 4 kilometres deep as well as TauTona, which is also part of the AngloGold Ashanti mines. It extends beyond 3.9 kilometres.
- Platinum: Total worldwide reserves of platinum metals are estimated to be some 69 000 metric tons. 95% is in the Earth's crust in South Africa. Platinum in most cases is a by-product of copper and nickel mines and refineries. The largest pure platinum mines are in South Africa, which is also the world's leading producer of the element. Around 120 metric tons were produced there in 2020. Bathopele mine is one of the largest platinum mines with reserves of up to 5.3 million ounces (oz) and a production of 120 000oz of platinum yearly. It is situated in Rustenburg. The largest platinum mining company worldwide is South Africa's Anglo American Platinum Ltd., headquartered in Johannesburg. Other leading platinum companies are Impala Platinum (South Africa), Sibanye Gold Ltd. (South Africa), and Lonmin (United Kingdom). Platinum is used for vehicle emission control devices and jewellery but is also in high demand for laboratory devices and as an investment.
- **Diamonds:** Diamond is probably one of the most valuable minerals in South Africa, being a valuable and majorly sought-after precious stone. South Africa's diamond production amounted to 7.2 million carats, positioning it as the world's sixth-largest diamond producer.
- **Coal:** South Africa is one of the top five coal exporting countries worldwide. Almost one quarter of the coal mined in the country gets exported making it the third largest foreign income earner for the state. The coal mine companies include South32's South Africa Energy Coal, Sasol Mining, Anglo American

PLC, Exxaro and Glencore Xstrata.



The <u>full solutions</u> are at the end of the unit

Cement

PPC and Lafarge are the two largest cement manufacturers in South Africa. PPC has 11 cement factories in Southern Africa. PPC's capacity is around 11.5 million tonnes of cement products each year and it has 26 batching plants across South Africa. PPC also produces aggregates with its Mooiplaas aggregates quarry in Gauteng having the largest aggregate production capacity in South Africa. Lafarge has 19 aggregate quarries in South Africa and manufactures cement in several locations.

Cement is a binder and is used for construction as it sets, hardens, and adheres to other materials to bind them together. Cement is seldom used on its own, but to bind sand and gravel (aggregate) together. Cement mixed with fine aggregate produces mortar for masonry, or with sand and gravel, produces concrete. Concrete is the most widely used material in existence and is behind only water as the planet's most-consumed resource.

Cements used in construction are usually inorganic, often lime or calcium silicate based, which can be characterised as non-hydraulic or hydraulic respectively, depending on the ability of the cement to set in the presence of water.

Non-hydraulic cement does not set in wet conditions or under water. Rather, it sets as it dries and reacts with carbon dioxide in the air. It is resistant to attack by chemicals after setting.

Hydraulic cements (e.g. Portland cement) set and become adhesive due to a chemical reaction between the dry ingredients and water. The chemical reaction results in mineral hydrates that are not very water-soluble and so are quite durable in water and safe from chemical attack. This allows setting in wet conditions or under water and further protects the hardened material from chemical attack.



Figure 8: Cement manufacture process

Cement is manufactured through a controlled chemical combination of calcium, silicon, aluminium, iron, and other ingredients. Common materials used to manufacture cement include limestone, shells, and chalk or marl combined with shale, clay, slate, blast furnace slag, silica sand, and iron ore. These ingredients, when heated at high temperatures form a rock-like substance that is ground into the fine powder that we commonly think of as cement. Concrete is formed when Portland cement creates a paste with water that binds with sand and rock to harden.

Note

To understand how cement is made you can visit the site <u>www.howcementismade.com</u> and interact with a 3-D cement plant.



Explosives

Explosives are also manufactured in South Africa. Explosives are used in underground mines and quarries to break coal and other rocks. An explosive is a solid or a liquid substance or mixture of substances which change themselves instantaneously into a large volume of gases at high temperature and pressure when a flame, heat, or sudden shock (detonation) is applied to it. Detonation is a process of giving a sufficiently violent shock to the explosive to bring about an almost instantaneous rearrangement of atoms.

The chemical energy in an explosive is released and converted suddenly into heat and mechanical energy when heat, flame or detonation is applied, and this conversion is due to a chemical reaction which is essentially a process of oxidation. An explosive contains enough oxygen necessary for complete oxidation.

Explosives are grouped in two types depending upon the speed with which the explosive effect is produced: low explosives and high explosives.



Figure 9: Gunpowder is a common example of a low explosive; and ammonium nitrate, nitro-glycerine, T.N.T., special gelatine, slurry explosives, are high explosives

When a low explosive is blasted, it is spread by rapid combustion and the effect of explosion is low. A low explosive is fired by ignition or a flame. High explosives always contain an ingredient which is explosive in itself, at least when sensitised by proper means. A high explosive explodes when a violent shock is applied to it with the help of a detonator; the explosion is instantaneous and reacts with high velocity. High explosives therefore produce a shattering effect.

Ammonium nitrate consumption for explosives has grown because of its safety advantage over other products such as dynamite. Ammonium nitrate can be shipped and stored and mixed with fuel oil when needed. Ammonium nitrate fuel oil (ANFO) is made of about 94% ammonium nitrate and 6% fuel oil. ANFO is widely used as an explosive in mining, quarrying, and tunnelling construction or wherever dry conditions exist.

AECI has been manufacturing explosives for mining in Modderfontein, Gauteng since 1895. AECI make blasting caps and detonation fuse wires specifically for the mining industry.

Fertiliser

Plants are only able to absorb nutrients from the soil when they are dissolved in water so that their root systems can absorb the nutrients. Nitrogen gas (N₂) for example, cannot be absorbed in the gas form, and needs to be changed into an ion that is soluble in water, for example the nitrate ion (NO_3^-). In the same way phosphorus is absorbed as phosphate ions (PO_4^3 -).

Fertilisers usually provide the three major plant nutrients (nitrogen, phosphorus, and potassium). Fertilisers are in general applied to the soil so that the nutrients are absorbed by plants through their roots. Often, cultivation practices do not give the soil enough time to recover and to replace the nutrients that have been lost naturally. Today, fertilisers play an important role in restoring soil nutrients so that crop yields can stay high. Some of these fertilisers are organic (e.g. compost, manure, and fishmeal). Industrial fertilisers are inorganic (for example ammonium nitrate (NH_4NO_3) or super phosphates ($Ca(H_2PO_4)_2$) and have the advantage of being in a soluble form that can be absorbed by a plant immediately.

Fertiliser packaging contains a set of numbers, for example 6 : 1 : 5. These numbers are called the NPK ratio, and they give the mass ratio of nitrogen, phosphorus, and potassium in the fertiliser. The NPK ratio expresses the content of each nutrient as a percentage of N, P and K in this order. A number in brackets after this ratio indicates the percentage by mass of N, P and K that is present in the fertiliser (what percentage of the total fertiliser is N, P and K). For example, in:

ΝΡΚ

$3\ 1\ 5\ (38)$

38% of the total fertiliser is nitrogen, phosphorus, or potassium
% N: 3 in every 9 parts of the 38% contains nitrogen (N)
% P: 1 in every 9 parts of the 38% contains phosphorus (P)
% K: 5 in every 9 parts of the 38% contains potassium (K)



Figure 10: Industrially made fertiliser indicating the NPK ratio
Depending on the types of plants you are growing, and the growth stage they are in, you may need to use a fertiliser with a slightly different ratio. For example, if you want to encourage root growth in your plant you might choose a fertiliser with a greater ratio of phosphorus in it. The main functions are summarised in the table below.

Nutrient	Promotes	When to use
Nitrogen (N)	leafy plant growth, faster plant growth	on lawns and other plants with lots of green leaves
Phosphorus (P)	strong roots, healthy fruit, blooming	on flowers and flower beds
Potassium (K)	disease resistance, growth of fruit	on fruit bearing plants

Some countries express the phosphorus content as P_2O_5 and potassium content as K_2O . South Africa expresses the NPK ratio in terms of the elements present as explained above. The rest of the fertiliser (62%) is made up of fillers, such as gypsum, lime, and sand. Other micronutrients, such as calcium (Ca), sulfur (S) and magnesium (Mg), are often added to the mixture.

The manufacture of fertilisers



Figure 11: The industrial manufacturing of fertilisers

Various chemicals are needed to make fertilisers, namely hydrogen, nitrogen and ammonia.

• Obtaining hydrogen: Fossil fuels are the main source of industrial hydrogen. Hydrogen can be generated from natural gas or coal. These processes are used by Sasol at their Gas-to-Liquid (GTL) and Coal-to-Liquid (CTL) facilities. Hydrogen is usually produced by the steam reforming of methane gas (natural gas). At high temperatures ($700^{\circ}C - 1\ 100^{\circ}C$), steam (H₂O) reacts with methane (CH₄) in an

endothermic reaction to yield syngas, a mixture of carbon monoxide (CO) and hydrogen (H_2) .

$$\rm CH_{4(g)} \! + \! H_2O_{(g)} \to \rm CO_{(g)} \! + \! 3H_{2(g)}$$

During a second stage, which takes place at a lower temperature of about 130° C, the exothermic reaction generates additional hydrogen. This is called a water gas shift reaction. Essentially, the oxygen (O) atom is stripped from the additional water (steam) to oxidise CO to CO₂. This oxidation also provides energy to maintain the reaction.

 Obtaining nitrogen: Fractional distillation is a separation method. It uses the difference in boiling temperatures of the components of a mixture to separate those components. A mixture is heated to convert the components into the vapour (gas) phase. The vapour mixture is then pumped into a tall separation column (called a fractional distillation column), usually at the bottom of the column. As the vapour mixture moves up the column and cools, the different components (called fractions) condense as the temperature drops below the various boiling point temperatures.

Air is a mixture of gases, mainly nitrogen and oxygen. Liquefied air (compressed and cooled to -200°C) is pumped into the fractional distillation column. Nitrogen gas has the lowest boiling point temperature and is collected at the top of the column.

 Producing ammonia and nitric acid: Ammonia (NH₃) plays an important role in the manufacturing process of fertilisers. The industrial process used to produce ammonia is called the Haber process. In this reaction nitrogen gas and hydrogen gas react to produce ammonia gas. The equation for the Haber process is:

 $N_{2(g)}\!+\!3H_{2(g)} \rightleftharpoons 2NH_{3(g)}$

This reaction takes place in the presence of an iron catalyst under high pressure (200 atmospheres (atm)) and temperature ($450^{\circ}C - 500^{\circ}C$) conditions.



Take note!

The forward reaction of the Haber process is exothermic, so the forward reaction is favoured by low temperatures. However, low temperatures slow all chemical reactions. So as the Haber process requires high temperatures, the ammonia is removed as soon as it is formed to prevent it being used in the reverse reaction.

The Ostwald process is used to produce nitric acid from ammonia. Nitric acid can then be used in reactions that produce fertilisers. Ammonia is converted to nitric acid in a three-step process.

Firstly ammonia is oxidised by heating it with oxygen, in the presence of a platinum (Pt) catalyst, to form nitrogen monoxide (NO) and water. This step is strongly exothermic, which makes it a useful heat source. The reaction that takes place is:

 $4NH_{3(g)}\!+\!5O_{2(g)}\rightarrow 4NO_{(g)}\!+\!6H_2O_{(g)}$

Secondly, nitrogen monoxide is oxidised again to yield nitrogen dioxide (NO_2) according to the following reaction:

$$2\mathrm{NO}_{(\mathrm{g})}\mathrm{+O}_{2(\mathrm{g})}
ightarrow 2\mathrm{NO}_{2(\mathrm{g})}$$

Thirdly, nitrogen dioxide is absorbed by water to produce nitric acid (HNO_3) as follows:

 $3\mathrm{NO}_{2(\mathrm{g})}\!+\!\mathrm{H}_{2}\mathrm{O}_{(\mathrm{l})}
ightarrow 2\mathrm{HNO}_{3(\mathrm{aq})}\!+\!\mathrm{NO}_{(\mathrm{g})}$

Nitrogen monoxide is a by-product of this reaction and is recycled, and the acid is concentrated to the required strength.

Nitric acid and ammonia can react together in an acid-base process to form the salt, ammonium

nitrate (NH₄NO₃). Ammonium nitrate is soluble in water and is often used in fertilisers. The reaction is as follows: $NH_{3(l)} + HNO_{3(l)} \rightarrow NH_4NO_{3(s)}$.

Ammonium nitrate is used to make fertilisers and explosives.

Note

To consolidate your understanding of how fertilisers work, watch this video by the agriculture academy called How fertilisers work.

How fertilisers work (Duration: 04.20)





Exercise 1.3

- 1. A bag of fertiliser has the following information on it: 3 1 6 (35). What percentage of the fertiliser is composed of nitrogen, phosphorus, and potassium?
- 2. Consider only the percentage of the fertiliser that is composed of nitrogen, phosphorus, and potassium. In a sample of 10 units, how many units are nitrogen, how many units are phosphorus and how many units are potassium?
- 3. Why does a plant need:
 - a. nitrogen
 - b. phosphorus
 - c. potassium
- 4. Which one of the following pairs of reactants is used in a reaction during the Haber process?
 - a. $\,H_{2(g)} \, \text{and} \, N_{2(g)}$
 - b. $SO_{3(g)}$ and $O_{2(g)}$
 - c. $NO_{(g)}$ and $O_{2(g)}$
 - d. $H_2O_{(l)}$ and $H_2S_2O_{7(l)}$

The <u>full solutions</u> are at the end of the unit.

The chloralkali Industry

The chlorine-alkali (chloralkali) industry is an important part of the chemical industry, which produces

chlorine and sodium hydroxide through the electrolysis of the raw material brine. Brine is a saturated solution of sodium chloride (NaCl) that is obtained from natural salt deposits.

Remember that electrolytic cells are used to transform reactants into products by turning electric current into chemical potential energy.

The products of the chloralkali industry have a number of important uses.

Chlorine is used:

- \cdot to purify water
- as a disinfectant.

Chlorine is used in the production of:

- hypochlorous acid (used to kill bacteria in drinking water)
- paper
- food
- antiseptics, insecticides, medicines, textiles, laboratory chemicals
- paints, petroleum products, solvents, plastics (such as polyvinyl chloride).

Chlorine is the chemical most often used to keep swimming pools and jacuzzis free of bacteria that can be hazardous to humans. The chlorine breaks down into many different chemicals, including hypochlorous acid (HOCl) and hypochlorite ions (OCl⁻). Both kill microorganisms and bacteria by attacking the cell walls and destroying the enzymes and structures inside the cell, making them harmless.

Sodium hydroxide (also known as 'caustic soda') is used to:

- make soap and other cleaning agents
- purify bauxite (the ore of aluminium)
- make paper
- make rayon (artificial silk).

One of the problems of producing chlorine and sodium hydroxide is that when they are produced together the chlorine combines with the sodium hydroxide to form chlorate (ClO^-) and chloride (Cl^-) ions. This leads to the production of sodium chlorate, NaClO, a component of household bleach.

To overcome this problem the chlorine and sodium hydroxide must be separated from each other so that they do not react. There are three industrial processes that have been designed to overcome this problem. All three methods involve electrolytic cells.

The mercury cell method

A mercury cell is an electrolytic cell.



Figure 12: A mercury cell

At the anode side, the anodes are placed in the aqueous NaCl solution, above the liquid mercury. The reduction of Cl^{-} occurs to produce chlorine gas, $Cl_{2(g)}$.

At the cathode side, a layer of $Hg_{(l)}$ at the bottom of the tank serves as the cathode. With a mercury cathode, the reaction of $H_2O_{(l)}$ to H_2 has a fairly high over potential, so the reduction of Na^+ to Na occurs instead. The Na is soluble in $Hg_{(l)}$ and the two combine to form the Na-Hg alloy amalgam. This amalgam can be removed and then mixed with water to cause the following reaction: $2Na~(in~Hg) + 2H_2O \rightarrow 2Na^+ + 2OH^- + H_{2(g)} + Hg_{(l)}$

The $Hg_{(l)}$ that forms, is recycled back into the liquid at the bottom of the tank that acts as a cathode. H₂ gas is released. NaOH is left in a very pure, aqueous form.

This method only produces a fraction of the chlorine and sodium hydroxide that is used by industry as it has certain disadvantages:

- mercury is expensive and toxic
- \cdot some mercury always escapes with the brine that has been used
- mercury reacts with the brine to form mercury(II) chloride
- the mercury cell requires a lot of electricity
- although the chlorine gas produced is very pure, mercury must be removed from the sodium hydroxide and hydrogen gas mixture. In the past the effluent was released into lakes and rivers, causing mercury to accumulate in fish and other animals feeding on the fish. Today, the brine is treated before it is discharged so that the environmental impact is lower.



Take note!

If chlorine comes into contact with hydrogen, it produces a mixture which will explode violently on exposure to sunlight or heat. Hydrogen chloride gas would be produced. Obviously, the two gases need to be kept apart.

The diaphragm cell method

The diaphragm cell is another electrolytic cell.





In a diaphragm cell:

- a porous diaphragm made of asbestos divides the electrolytic cell into an anode compartment and a cathode compartment
- brine is introduced into the anode compartment and flows through the diaphragm into the cathode compartment
- an electric current is passed through the brine causing the salt's chlorine ions and sodium ions to move to the electrodes
- + Chlorine gas is produced at the anode: $2 \mathrm{Cl}_{\mathrm{(aq)}} + 2 \mathrm{e}^{-}
 ightarrow \mathrm{Cl}_{2(\mathrm{g})}$
- At the cathode, sodium ions react with water forming caustic soda (NaOH) and hydrogen gas: $2Na^+{}_{(aq)}+ 2H_2O_{(l)}+e^- \rightarrow 2NaOH_{(aq)}+H_{2(g)}$
- Some NaCl salt remains in the solution with the caustic soda and can be removed at a later stage.

The advantages of the diaphragm cell are:

- it uses less energy than a mercury cell
- it does not contain toxic mercury.

It also has disadvantages however:

- \cdot the sodium hydroxide is much less concentrated and not as pure
- \cdot the chlorine gas often contains oxygen gas as well
- the process is less cost-effective as the sodium hydroxide solution needs to be concentrated and purified before it can be used
- asbestos poses health risks to workers.

The membrane cell method

A third way to make even more pure N_aOH is to use a membrane cell. It is preferred over the diaphragm cell or mercury cell methods because it uses the least amount of electric energy and produces the highest quality N_aOH . An ion-permeable membrane is used to separate the anode and cathode.



Figure 14: A membrane cell

At the anode, saturated brine is fed to the compartment containing $_{Na^+}$ and $_{Cl^-}$. Chloride ions are oxidised to chlorine gas at the anode.

$2\mathrm{Cl}_{\mathrm{(aq)}}+~2\mathrm{e}^{-} ightarrow\mathrm{Cl}_{2(\mathrm{g})}$

The chlorine is purified by liquifying it under pressure. The oxygen stays as a gas when it is compressed at ordinary temperatures. The membrane allows Na^+ ions to pass into the cathode compartment.

At the cathode, ${
m H_2O}$ is reduced to form ${
m OH}^{-}$ and ${
m H_2}$ gas: ${
m 2H_2O}
ightarrow {
m H_2+2OH}^{-}$.

 Na^+ ions that had flowed through the membrane and OH^- produced through the reduction of water react to form aqueous NaOH. H₂ gas also is produced as a by-product.

The membrane is made from a polymer which only allows positive ions to pass through it. That means that the only the sodium ions from the sodium chloride solution can pass through the membrane, and not the chloride ions. The advantage of this is that the sodium hydroxide solution being formed in the right-hand compartment never gets contaminated with any sodium chloride solution. The sodium chloride solution being used must be pure. If it contains any other metal ions, these would also pass through the membrane and so contaminate the sodium hydroxide solution.

 $2NaCl_{(aq)} + 2H_2O_{(l)} \rightarrow Cl_{2(g)} + H_{2(g)} + 2NaOH_{(g)}$

The advantages of using this method are:

- the sodium hydroxide that is produced is very pure because chloride ions cannot pass through the membrane and contaminate it
- the sodium hydroxide has a relatively high concentration
- this process uses the least electricity of all three cells
- the cell is cheaper to operate than the other two cells
- the cell does not contain toxic mercury or asbestos.

Note

To further understand the chloralkali industry, watch this video by Fuse Schools, called The Brine Industry.

The Brine Industry (Duration: 04.57)





Exercise 1.4

1. Summarise what you have learnt about the three types of cells in the chloralkali industry by copying and completing the table below:

	Mercury cell	Diaphragm cell	Membrane cell
Main raw material			
Mechanism of separating where Cl_2 and NaOH are produced			
Anode reaction			
Cathode reaction			
Purity of NaOH produced			
Purity of Cl_2 produced			
Energy consumption			
Environmental impact			
Cost of production			

2. The diagram below shows the sequence of steps that take place in a mercury cell.



Summary

In this unit you have learnt the following:

- The growth of South Africa's chemical industry was largely due to the mining industry, which requires explosives for their operation. One of South Africa's major chemical companies is Sasol. Other important examples of chemical industries in the country are the chloralkali and fertiliser industries.
- The Fisher-Tröpsch process is used by Sasol to convert coal to liquid fuels
- Other products made from this process include heavy waxes, organic liquids, and chemical feedstocks.
- South Africa has many mineral resources which are mined from the ground.
- These mined minerals include iron ore, coal, gold, platinum, and diamonds.
- Iron ore is processed in blast furnaces to produce pure iron which is used to make steel.
- South Africa has a large paint, cement, and adhesive industry.
- The fertiliser industry is very important in providing fertilisers with the correct nutrients in the correct quantities to ensure maximum growth for various plants and crops.
- Fertilisers can be produced industrially using several chemical processes: the Haber process reacts nitrogen and hydrogen to produce ammonia; the Ostwald process reacts oxygen and ammonia to produce nitric acid.
- Ammonium nitrate produced from the Ostwald process is used by companies like AECI to manufacture explosives.
- The chloralkali industry produces sodium hydroxide and chlorine.
- Chlorine is used to make paper, as a disinfectant.
- Sodium hydroxide is used to make soap and paper.
- The production of chlorine and sodium hydroxide is done in electrolytic cells: a mercury cell, a diaphragm cell, or a membrane cell.

Unit 1: Assessment

Suggested time to complete: 20 minutes

- 1. A bag of fertiliser has the following information on it: 18 21 6 (35) and has a mass of 3 kg. Calculate the mass of nitrogen (N), phosphorus (P) and potassium (K) that is present in the bag of fertiliser.
- 2. You have determined that your vegetable garden will need 12 g of nitrogen. The fertiliser you have has 4:2:3 (40) written on the packet. How many grams of fertiliser will you need for your garden?
- 3. The reaction of hydrogen and nitrogen is an exothermic reversible reaction.
 - a. Write a balanced equation for this reaction.
 Use your knowledge of chemical equilibrium and explain what effect the following will have on this equilibrium reaction:
 - b. raising the temperature
 - c. raising the pressure.
- 4. Refer to the flow diagram, which shows the reactions that take place in a membrane cell, and then answer the questions that follow.



- a. What liquid is present in the cathode compartment at A?
- b. Identify the gas that is produced at B.
- c. Explain one feature of this cell that allows the $_{Na^{+}}$ and $_{OH^{-}}$ ions to react at C.
- d. Give a balanced equation for the reaction that takes place at C.
- 5. Approximately 30 million tonnes of chlorine are used throughout the world annually. Chlorine is produced industrially by the electrolysis of brine. The diagram represents a membrane cell used in the production of Cl_2 gas.



- a. What ions are present in the electrolyte in the left-hand compartment of the cell?
- b. Give the equation for the reaction that takes place at the anode.
- c. Give the equation for the reaction that takes place at the cathode and forms a gas.
- d. What ion passes through the membrane while these reactions are taking place?
- e. Chlorine is used to purify drinking water and swimming pool water. The substance responsible for this process is the weak acid, hypochlorous acid (HOCl). One way of putting HOCl into a pool is to bubble chlorine gas through the water. Give an equation showing how bubbling $Cl_{2(g)}$ through water produces HOCl.

The <u>full solutions</u> are at the end of the unit.

Unit 1: Solutions

Exercise 1.1

1.

- a. Coal is first converted to gas and then its hydrogenation takes place.
- b. An iron catalyst because the coal/gas is of a low quality.
- 2. c. The gasoline fraction produced in Fischer-Tröpsch process is collected for further processing, while the heavy oil fractions are subjected to cracking to produce more gasoline.

Back to Exercise 1.1

Exercise 1.2

- 1. To make steel, as a catalyst in the Fischer-Tröpsch process.
- 2. Gold, platinum, and diamonds.
- 3. Gold reserves are running out/gold reserves are too deep to make mining them viable.

Back to Exercise 1.2

Exercise 1.3

1. The number in brackets is the percentage of the fertiliser that contains nitrogen, phosphorus, and potassium:

N:P:K (percentage) $3\ -\ 1\ -\ 6\ (35)$ In this example the percentage is 35%

2. Information about nitrogen, phosphorus and potassium is given for fertilisers as: N:P:K (percentage) 3 - 1 - 6 (35) The ratio of N:P:K in this fertiliser is 3:1:6. Therefore in 10 units of the fertiliser:

3 units of the 35% are nitrogen 1 unit of the 35% is phosphorus 6 units of the 35% are potassium

Notice that these three numbers add up to 10.

- 3.
- a. leafy plant growth, faster plant growth
- b. strong roots, healthy fruit, blooming
- c. disease resistance, growth of fruit
- 4. a

Back to Exercise 1.3

Exercise 1.4

1.

	Mercury cell	Diaphragm cell	Membrane cell
Main raw material	brine	brine	brine
Mechanism of separating	reduction of Na^+ to Na in Hg amalgam which flows to separate compartment	use of a porous asbestos diaphragm	use of a membrane (charge specific)
Anode reaction	$\mathrm{Cl}^{\text{-}} ightarrow \mathrm{Cl}_{2\mathrm{(g)}}$ + 2e-	$\mathrm{Cl}^{\text{-}} ightarrow \mathrm{Cl}_{2\mathrm{(g)}}$	$\mathrm{Cl}^{\text{-}} \to \mathrm{Cl}_{2(g)}$
Cathode reaction	$2H_2O_{(l)}\!+\!2e^{-}\to H_{2(g)}\!+\!2OH^{-}{}_{(aq)}$	$2H_2O_{(l)}{+}2e^{-}\rightarrow H_{2(g)}{+}2OH^{-}_{(aq)}$	$2H_2O_{(l)}\!+\!2e^{-}\to H_{2(g)}\!+\!2OH^{-}{}_{(aq)}$
Purity of NaOH produced	low purity	medium purity, low concentration	high purity
Purity of Cl_2	high purity	mixed with O_2	mixed with O_2
Energy consumption	high	less than the mercury cell	low energy
Environmental impact	high	lower than the mercury cell, asbestos is a health hazard	ower than the diaphragm and mercury cell, mercury is toxic
Cost of production	expensive	expensive, as product must be purified	low production cost

2.

- a. aqueous sodium chloride
- b. $2Cl_{(aq)} + 2e \rightarrow Cl_{2(g)}$
- c. sodium mercury amalgam

d. sodium mercury amalgam and water

Back to Exercise 1.4

Unit 1: Assessment

1. $T\frac{27}{0.40} = 67.5$ g.The ratio of N:P:K in this fertiliser is 18:21:6.

35% of the sample contains nutrients. We need to convert this percent to mass as follows: 35% of 3 kg = (0.35) (3) = 1.05 kg. Therefore 1.05kg of the fertiliser contains nutrients. For every 18 units of nitrogen there are 21 units of phosphorus and 6 units of potassium so the total number of units is 45.

18 of the 45 units are nitrogen, therefore: $rac{18}{45} imes 1.05 = 0.42~{
m kg}$ will be nitrogen.

21 of the 45 units are phosphorus, therefore: $rac{21}{45} imes 1.05 = 0.49~
m kg$ will be phosphorus.

6 of the 45units are potassium, therefore: $rac{6}{45} imes 1.05 = 0.14~{
m kg}$ will be potassium.

2. The NPK mass ratio is $4 : 2 : 3 (40) \cdot 4 + 2 + 3 = 9$.

 $4~{
m of}$ the $9~{
m units}$ are nitrogen and $12~{
m g}$ of nitrogen is needed, therefore:

 $\frac{4}{9}$ ×total mass of nutrients required = 12 g.

Total mass of nutrients required = 12 g $\times \frac{9}{4}$ = 27 g

40% of the sample contains nutrients, therefore: 40% of fertiliser required = 27 g Total fertiliser required = $\frac{27}{0.40} = 67.5$ g

3.

- a. $N_{2(g)} \! + \! 3H_{2(g)} \rightleftharpoons 2NH_{3(g)}$
- b. When the temperature of any reaction is increased, the rate of both the forward and the reverse reactions will be increased. However, the forward reaction is exothermic. Raising the temperature would favour the reaction which would remove the excess heat. The reverse reaction will therefore be favoured more than the forward reaction. This will result in more $\rm NH_3$ decomposing to form $\rm H_2$ and $\rm N_2$. The overall result will be that more reactants will be formed.
- c. Raising the pressure will shift the equilibrium to decrease the pressure. This would favour the reaction which produced fewer gas molecules. The forward reaction produces 2 moles of gas (compared to the 4 moles of the reverse reaction). The forward reaction will therefore be favoured. The overall result will be that more products will be formed.

4.

- a. Water and dilute $\rm NaOH$
- b. chlorine gas
- c. This type of cell has an ion permeable membrane that allows the positive $_{Na^+}$ ions to move from the anode compartment to the cathode compartment and interact with the OH⁻ ions.

d.
$$\operatorname{Na^{+}}_{(aq)} + OH^{-}_{(aq)} \rightarrow \operatorname{NaOH}_{(aq)}$$

5.

- a. chloride ions (Cl) and sodium ions (Na^+)
- b. $2\mathrm{Cl}_{(\mathrm{aq})} + 2e^{-} \rightarrow \mathrm{Cl}_{2(\mathrm{g})}$
- c. $2H_2O_{(l)} + 2e^- \rightarrow H_{2(g)} + 2OH^-_{(aq)}$
- d. sodium ions

e. $Cl_{2(g)}+H_2O_{(l)} \rightarrow HOCl_{(aq)}+HCl_{(aq)}$

Back to Unit 1: Assessment

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