TOPIC 1: INTRODUCTION TO PHYSICS

Physics falls under a broader category of Science. Science is divided into three areas namely Biology, Physics and Chemistry. The main objective of these subjects is to study and try to understand the universe and everything in it. The three subjects are very much interrelated.

Specifically, Physics is a branch of science that deals with the study of matter and how it relates to energy. *What is matter?* Matter is everything that we see and interact with which has mass and occupies some space.

In trying to study the universe, Physicists in most cases endeavor to carry out experiments where quantities are measured and then collected as data. Such data and observations made are then used to study any logical pattern(s) to reveal any relationship between the data. Before the actual experimentation, propositions can be made which are actually investigated by the experiment. This is called a **hypothesis**. The data is taken through a process of analysis and interpretation and conclusions drawn. The conclusion may or may not concur with the hypothesis earlier stated. In case of any pattern between the quantities, this can be expressed in the form of a **law, principle or theory**.

A law is a description of a relationship between quantities that manifests itself in recurring patterns of events. Some of the laws we will be looking at include Hooke's law, Ohm's law, etc.

A theory is an explanation of phenomena in terms of most basic natural processes and relationships. Theories are tentative but can help us affirm already known laws. When tested and proved, a theory can become a law. A theory that has survived longer time is assumed to possess some measure of truth.

Physics as a subject is divided into six broad branches as discussed below:

i. **Mechanics**

This branch deals majorly with motions under the influence of forces. Under this branch, we look into details the aspects of linear, circular and oscillatory motions as well as motion of fluids.

ii. **Optics**

This branch takes a keen look at the behavior of light in various media.

iii. **Electricity and magnetism**

This branch looks at the interaction between electric fields and magnetic fields and the applications of such interactions.

iv. Thermodynamics

This branch looks at how heat as a form of energy is transformed to/from other forms of energy.

v. Atomic Physics

This area of study is targeted at the behavior of particles of the nucleus and the accompanying energy changes.

vi. Waves

It deals with the study of the propagation of energy through space.

Physics does not only relate the remaining two science subjects but also enjoys a relationship with other subjects as well. For instance, it is the foundation of **technological development**in any country.

The study of Physics can open up many avenues of professions including engineering, degree, diploma or certificate courses.

N/B. Add the relationship between physics and other disciplines,career opportunities in physics and basic laboratory rules.

TOPIC 2: MEASUREMENTS

2.1: Physical quantities

The study of physics deals mostly with physical quantities. It involves investigation, making observations, collection of data and their analysis, interpretation and drawing conclusions.one important aspect in data collection is taking measurements. When taking measurements, one must first know the quantity to be measured and the unit for measuring it.

For purposes of uniformity, physical quantities have been assigned specific units which are acceptable world over. These are referred to as **International Standard Unit (**SI Unit).

Physical quantities are categorised into two namely basic physical quantities and derived physical quantities.

A basic physical quantity is a quantity that cannot be obtained by either multiplication or division of other basic physical quantities. They are seven in number. The table below shows the seven basic physical quantities, their symbols, SI units and the symbols to the units:

Basic physical quantity	Symbol of quantity	SI Unit	Symbol of unit
Length	L	Metre	M
Mass	M	Kilogram	Kg
Time	T	Second	S
Electric current	I	ampere	A
Thermodynamic temperature	T	Kelvin	K
Amount of substance		Mole	Mol
Luminous intensity		Candela	Cd

A derived physical quantity is one which can be obtained by either multiplication or division of other basic physical quantities. The table shows some of the derived physical quantities, how they are obtained and their SI Units:

Derived physical quantity	How it is calculated	SI Unit	Symbol of unit
Area	Length*length	Square metre	m^2
Volume	Length*length*length	Cubic metre	m^3
Pressure	Force/area	newton per square metre or pascal	N/m^2 or Pa
Density	Mass/ volume	Kilogram per cubic metre	Kg/m^3

2.2: Length

Length can be defined as the distance between any two points. The SI unit of length is the metre (m). Other multiples and submultiples of the metre are stated below:

Kilometre (km); 1m= 10^{-3}km **Centimetre (cm); 1m= 100cm**

Hectometre (Hm); 1m= 10^{-2}Hm **Millimetre (mm); 1m= 1000mm**

Decametre (Dm); 1m= 10^{-1}Dm **Micrometre (μm); 1m= 10^{6}μm**

Decimetre (dm); 1m= 10dm

Length can be estimated or measured accurately using appropriate measuring instrument. The type of instrument to be used at any time depends on two factors:

- The object to be measured.
- The desired accuracy.

2.2.1: Estimation

This method involves comparing the object to be measured with another of standard measure. For example, the height of a tall flag post can be compared with that of a wooden rod whose length is known. Thus at any given time;

Height of flag post/ height of rod = length of shadow of post/ length of shadow of rod

From this expression, the height of the flag post can be estimated. Suppose the height of the rod= 1m, length of shadow of rod= 120cm and length of shadow of post= 480cm, then the height of the flag post is given by;

H_p/ 100cm =480cm/120cm

H_p= 100 x 4 = 400cm

Also, the thickness of a sheet of paper may be estimated by taking several sheets of the paper and measuring their thickness then dividing by the number of sheets of paper;

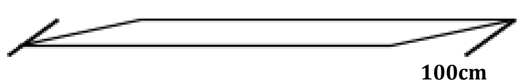
Thickness of a sheet of paper= thickness of n papers/ number of papers, n.

2.2.2: Accurate measurement

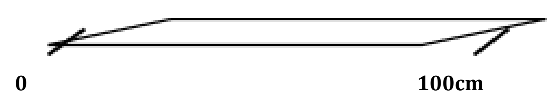
To measure length accurately, the following instruments can be used:

a) Metre rule

A metre rule is marked in centimetres. It is marked 0 and 100cm at its extreme ends. If some space is left before 0cm and 100cm mark then it is referred to as a ruler.



(a) Metre rule



(b) Ruler

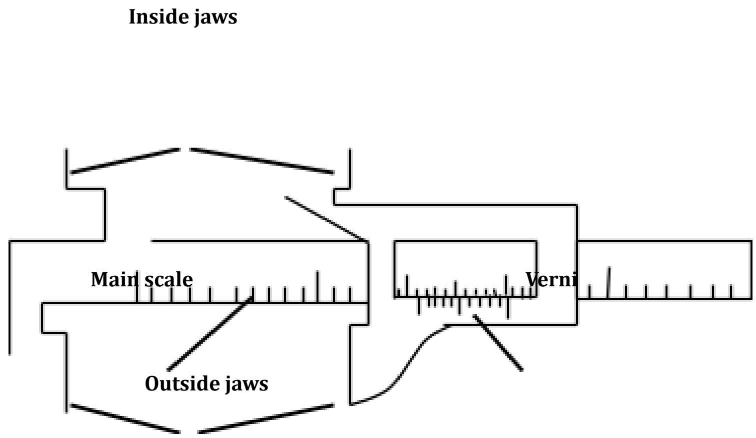
The smallest scale division of a metre rule is 0.1cm (1mm). The smallest scale division of any instrument is known as its accuracy. Thus the accuracy of a metre rule is 0.1cm.

When using a metre, one must ensure the following:

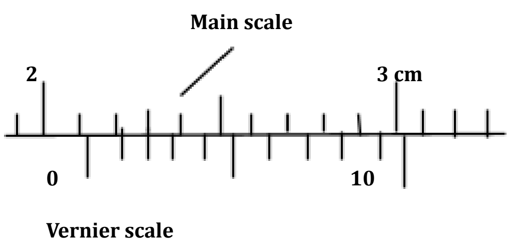
- That the object to be measured is in contact with the metre rule.
- That one end of the object is at 0cm mark.
- That the eye is perpendicular to the scale so as to avoid parallax error.

<u>b) Vernier calipers(form two work)</u>

A vernier calliper is more accurate compared to a metre rule. It has two scales; the main scale and vernier scale. It has an accuracy of 0.01cm compared to that of a metre rule of 0.1cm. It also has two jaws; the inside jaws and outside jaws. The object whose length is to be measured is placed between the outside jaws.



Consider a section of the scale shown below:



The smallest scale division of the main scale is 0.1cm. The whole of the vernier scale is of length 9mm (0.9cm) which is sub-divided into ten equal divisions. Therefore each division of the vernier scale represents 0.09cm. The accuracy of the callipers is given by the difference between the two least readings;

i.e. accuracy of a vernier calliper = 0.1 – 0.09 = 0.01cm

The reading by the vernier callipers can be obtained in three steps:

Step 1: Reading of the main scale. Take the reading of the main scale just before the zero mark of the vernier scale i.e. main scale reading= 2.1cm.

Step 2: Reading of the vernier scale. Check which mark on the vernier scale coincides exactly with a mark on the main scale i.e 2^{nd} mark. Vernier scale reading= (n^{th} mark x 0.01) cm.

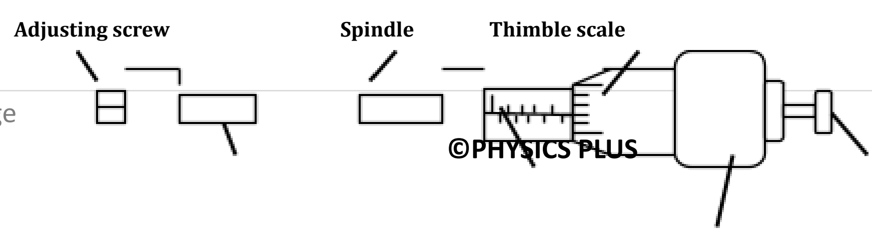
Step 3: Adding the two readings.

Therefore for the figure above, the reading= 2.1 + (2 x 0.01)

= 2.1+0.02

= 2.12cm

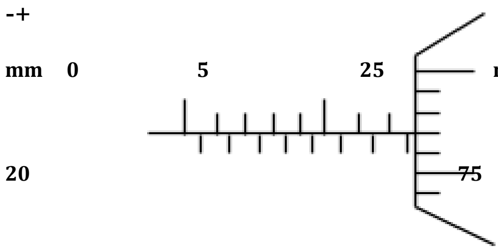
c) Micrometer screw gauge

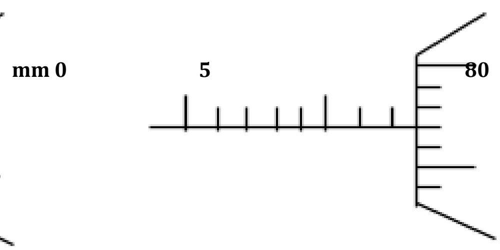


Anvil **Sleeve scale** **Ratchet**

Thimble

A micrometer screw gauge can be used to measure very small distances accurately like measuring the diameter of a thin wire. The accuracy of a micrometer screw gauge is 0.01mm. It also has two scales; sleeve scale and a thimble scale. The sleeve scale is marked in millimetres while the thimble scale is divided into either 50 or 100 equal divisions.





(a) Thimble having 50 divisions (b) Thimble having 100 divisions

The distance moved by the spindle in one complete rotation of the thimble is called a **pitch**.

In (a), the spindle advances or retreats by 0.5mm per complete rotation of the thimble;

50 divisions on the thimble scale= 0.5mm

And 1 division= 0.5mm/50 = 0.01mm

In (b), the spindle moves through 1mm per complete rotation of the thimble.

Similarly, 100 divisions= 1mm

Hence 1 division= 1mm/100 = 0.01mm

So whether the thimble has 50 or 100 divisions, the least reading of the micrometer screw gauge remains the same i.e 0.01mm.

Taking measurements using the micrometer screw gauge also involves three steps:

Step 1: Taking the sleeve scale reading. Read the observable mark at the edge of the thimble in mm.

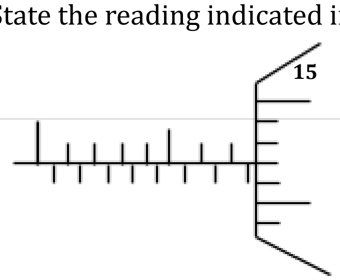
Step 2: Taking the thimble scale reading. Read off the mark on the thimble scale that coincides directly with the centre line of the sleeve scale.

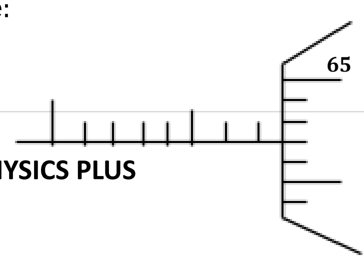
Hence the thimble scale reading= (n^{th} mark * 0.01) mm.

Step 3: adding the two readings. The sum of the two readings gives the reading by the micrometer screw gauge.

Example 2.1

1. State the reading indicated in each case:





0 5 5 10

10 60

(a) (b)

a) Reading =7.5 + (12 x 0.01)

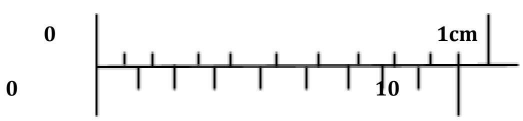
= 7.5+0.12 = 7.62mm

b) Reading = 12 + (62 x 0.01)

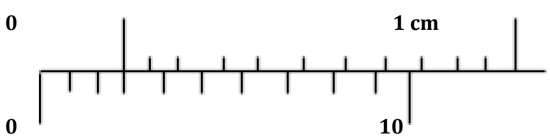
= 12+0.62 = 12.62mm

2.2.3: Zero error of a vernier calliper and micrometer screw gauge

When the jaws of the vernier calliper are closed without any object between them then the zero marks of the main scale and that of the vernier scale should coincide. In this case the calliper is said to have no zero error.

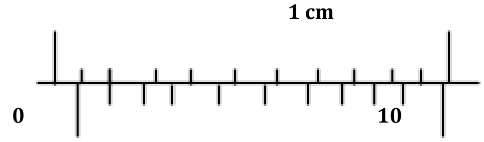


However, when the two zero marks do not coincide then the calliper has a zero error. When the zero of the vernier scale is found to the left of the zero of the main scale, the calliper is said to have a **negative zero error**. If the instrument is used with this error, then the reading obtained will be lower than the actual value. To get the actual reading the error must be added to the instrument's reading.



Zero error= - (3 x 0.01) = -0.03cm

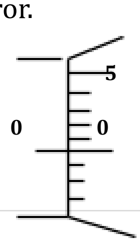
If the zero of the vernier scale is to the right of the zero of the main scale, the calliper has a positive zero error.

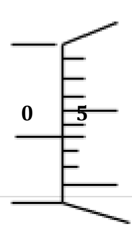


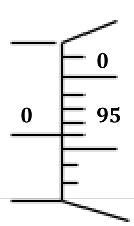
Zero error= + {0 + (1 x 0.01)} = +0.01cm

If the I calliper is used with such an error, the reading obtained will be higher than the actual value. Hence the error should be subtracted from the reading to get the correct value.

Similarly, when the micrometer screw gauge is closed without any object between its anvil and spindle and the 0mark of the thimble scale fails to coincide with the centre line of the sleeve scale then it is said to have a zero error. Note that the edge of the thimble should also be in line with the zero mark of the sleeve scale when taking the zero error.







0

(a)No error **(b) A negative zero error** **(c) A positive zero error**

In (b), the error= - (3 x 0.01) = -0.03mm

In (c), the error = + (4 x 0.01) = +0.04mm.

Generally, when the signs of the zero errors are taken into account, all zero errors are subtracted from the instrument's reading to give the correct reading.

Example 2.2

1. A micrometer screw gauge was used to measure the diameter of a thin wire. The reading is as shown in the figure below:

State the diameter of the wire if the micrometer had:

a) A negative zero error of 0.22mm

Correct reading = instrument's reading – zero error

= 7.52-(-0.22)

= 7.52+0.22 = 7.74mm

b) A positive zero error of 0.10mm

Correct reading= 7.52-(+0.10)

= 7.52-0.10 = 7.42mm

2.3: Area (form one)

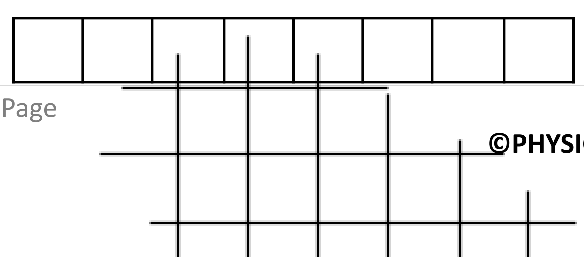
Area is defined as the measure of surface. Its SI Unit is the square metre (m^2). Other units of area include cm^2, km^2, hectares etc. area can also be estimated or calculated accurately. The area of regular shapes can be calculated from known formulas;

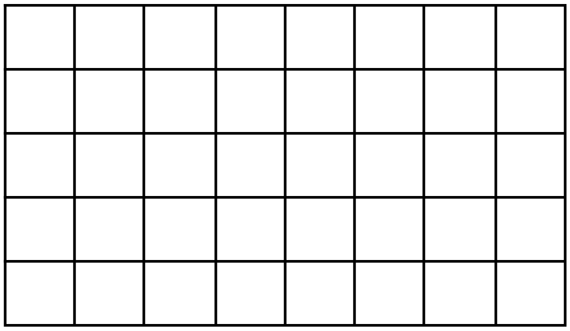
e.g. area of a triangle= ½(base x height)

Area of a rectangle= length x width

Area of a circle= πr^2

The area of irregular shapes can be estimated by counting the number of small squares which are covered by the irregular object. Not that in this case the area of each small must be known. Consider the figure below of an irregularly- shaped object.





The number of complete squares covered by the shape= 14

The number of incomplete squares covered by the shape=19

Therefore, the number of complete squares covered by the shape is approximately (14+ 19/2) = 23.5 squares.

Suppose the area of one square is $1cm^2$, then the area of the shape is approximately;

Area= 23.5 x 1 = $22cm^2$.

2.4: Volume

Volume is the amount of space occupied by an object. It's measured in cubic metre (m^3). Other commonly used units include cm^3, ml, litre etc.

2.4.1: Volume of regularly shaped objects

Volumes of regularly shaped objects can be calculated from known formulas. Generally, the volume a regular object is given the product of its cross section area by the height.

Volume = cross section area x height

2.4.2: Volume of liquids

Volumes of liquids can be measured using specific instruments. These include the measuring cylinder, volumetric flask, beaker, burette and pipette. Note that liquids have no definite shapes but take the shape of the container in which they are put.

2.4.3: Volume of irregularly shaped objects

Volumes of irregularly shaped solids can be obtained by displacement method. This may involve using a measuring cylinder or a Eureka can.

Using a measuring cylinder.

The cylinder is first filled with water upto a certain level and its volume read off. Then the object whose volume is required is immersed in the water completely. The new level is read off. The volume of the object is equal to the difference between the two readings.

Using the eureka can

The can is first filled with water until it overflows through the spout. After the flow has stopped, carefully immerse the object and collect the water which flows out through the spout using a measuring cylinder. Wait until the last drop comes out and then read off the volume of the water collected. This is equal to the volume of the solid immersed since it has been displaced by the solid.

These two methods only work when the solid:

- Does not react with water.
- Does not absorb water.
- Can sink in water (denser than water).

2.4.4: Volume of a floating object

The same method of displacement is used but with a sinker. It involves three steps:

1. Fill the eureka can with water.
2. Carefully immerse the sinker and note the volume of water it displaces.
3. Fill the eureka can again and now immerse both the sinker and the floating object. Note the new volume of water displaced.

The difference between the two volumes gives the volume of the floating object.

2.5: Mass

Mass is the quantity of matter in a body. It is measured in kilogram. Other commonly used units include gram, milligram and tonne.

1kg= 0.001ton

1kg= 1000g

1kg= 1000000mg

Mass of a body depends on the number of particles it has and its size. Hence the mass of a body is the same everywhere since the number of particles in a body is always constant. Mass is measured using a beam balance. There are other types of balances which may be used to measure mass. These include the top pan balance and a lever balance.

2.6: Density

Density is defined as the mass per unit volume of a substance;

Density= mass/volume

The SI Unit of density is the kilogram per cubic metre. The other commonly used unit is gram per cubic centimetre (g/cm^{-3}). The symbol of density is rho (ρ).

The density of a regularly shaped object can be obtained by first finding its mass and volume separately and then substituting in the formula, $\rho = m/v$.

Example 2.3

1. An object of mass 50.1g has a density of $16.7gcm^{-3}$. What is the volume of the object?

$\rho = m/v$

Therefore, $v = m/\rho = 50.1g/16.7gcm^{-3}$

$= 3cm^3$

2.6.1: Density bottle

This is a special glass bottle that can be used to determine the density of liquids and certain solids like lead shot. The bottle has its capacity indicated on its surface.

a) Density of a liquid using a density bottle

The empty bottle with its stopper is first weighed and its mass noted, say m_1. It is then filled with the liquid, stopper replaced and wiped carefully. It should always be held by its neck to avoid unnecessary heating and expansion of the bottle. The bottle is then weighed and the mass noted, say m_2.

(a) Empty bottle **(b) Bottle filled with water**

The difference between the two volumes is equal to the volume of the liquid in the bottle. Suppose the capacity of the bottle is V, then the density of the liquid is given by:

Density of liquid= $(m_2-m_1)/V$.

b) Density of a solid (lead shot) using a density bottle.

The empty bottle is first weighed and its mass noted; m_1. Some lead shot are added into the bottle and weighed again. Let the mass of the bottle and lead shot be m_2. Hence the mass of lead shot is (m_2-m_1). Water is then added into the bottle until it is completely filled. The new mass of the bottle and its contents noted. Let the mass of bottle, lead shot and added water be m_3. Thus the mass of water added is (m_3-m_2).

Since the density of water is $1gcm^{-3}$, the volume of water added to the bottle is given by;

$V_w = m/\rho = (m_3-m_2)/1 = (m_3-m_2)$.

Suppose the capacity of the bottle is V, then the volume of the lead shot is given by;

$V_L = V - V_w = V - (m_3-m_2)$.

And the density of lead shot = $(m_2-m_1)/\{V - (m_3-m_2)\}$.

Note:

1. There should be no air bubbles in the liquid being used to fill the bottle.
2. The solid being used in (b) above should be one that does not dissolve in and react with water.

Example2. 4

1. The mass a $50cm^3$ density bottle is 10.0g when empty and 60.0g when filled with copper turnings upto a certain level. Some water is added into the bottle until it is full. The mass of the bottle and its contents is found to be 90.0g. Determine the density of the copper turnings.

Mass of copper turnings= 60.0 – 10.0= 50.0g

Mass of water added= 90.0 – 60.0= 30.0g

Since ρ of water= $1gcm^{-3}$

The volume of water added= 30.0/1 = $30cm^3$

Therefore, the volume copper turnings=50 – 30= $20cm^3$.

Hence density of copper turnings=$50g/20cm^3$= $2.5gcm^{-3}$.

2.6.2: Density of mixtures

The density of mixture is the mass of the mixture divided by its volume;

Density of mixture= mass of mixture/ volume of mixture.

The density of a mixture always lies between the densities of its constituent substances.

Example 2.5

1. $1800cm^3$ of fresh water of density $1000kgm^{-3}$ is mixed with $2200cm^3$ of sea water of density $1025kgm^{-3}$. Calculate the density of the mixture in gcm^{-3}.

Volume of mixture= 1800 + 2200= $4000cm^3$

Mass of fresh water= 1800 x 1= 1800g

Mass of sea water = 2200 x 1.025= 2255g

Hence density of mixture = (1800+2255) g/ (1800+2200) cm^3

= 4055/4000 = $1.01375gcm^{-3}$

2.7: Time

Time is the measure of duration of an event. The SI Unit of time is the second (s). time can also be measured in microseconds (μs), milliseconds (ms), minutes (min), hours (hr), days, weeks, months, years etc. time can be measured using stop watches or stop clocks. A stop watch is more accurate compared to stop clock.

2.8: The oil drop experiment(form two)

The aim of this experiment is to estimate the size of an oil molecule. When a drop of oil is placed on the surface of clean water, it spreads out into a uniform circular patch. The oil lowers the surface tension of water whose particles then pull away from the oil.

The patch is assumed to have a thickness equal to the thickness of the oil molecule. The oil drop is also assumed to be a perfect sphere. Thus its volume is given by $4/3(\pi r^3)$. For clarity of the patch, some lycopodium powder is gently sprinkled of the surface of water.

So, volume of an oil drop= volume of patch.

$4/3(\pi r^3) = \pi r_p^2 t$;

Where r= radius of the oil drop,

r_p= radius of patch.

t= thickness of the oil molecule.

In the experiment, a number of assumptions are made. These include:

- Volume of an oil drop is equal to the volume of the patch.
- The oil drop is a perfect sphere.
- The patch is perfectly circular and one molecule thick.
- The oil drop is one molecule thick.

The knowledge of the oil drop experiment can be used to determine the extent of environmental damage as a result of oil spillage from ships in large water bodies.

TOPIC 3: <u>FORCE</u>

Force is defined as a push or a pull on a body.si unit of force is the newton (N).force is a vector quantity i.e. has both magnitude (size) and direction.

3.1: Effects of force on a body

- ✔ Sets a body into motion.
- ✔ Can stop a moving body.
- ✔ Can increase or reduce the speed of a moving body.
- ✔ Can change the direction of a moving body.
- ✔ May deform (change the shape of) a body.

Force is represented by a straight line with an arrow, which shows the direction in which it acts. One newton is the force which gives a mass of 1 kg an acceleration of 1 m/s^2.

3.2: Types of force

Force can be categorized in two ways. These are:

- As either a push or a pull
- As either contact or non-contact force

Contact forces are those forces between bodies which are in contact e.g. action and reaction, viscous drag, friction etc. Non-contact forces act between bodies at a distance e.g. gravitational force, magnetic force, electrostatic force etc.

Below are some common types of forces:

3.2.1: Gravitational force

It is the force that attracts all objects towards the centre of the earth. It is unique in every planet. The earth's gravitational force is that force of attraction between the earth and the body. The pull of gravity on a body towards the centre of the earth is called its weight. Weight varies from place to place although its mass remains constant.

The weight of a body is the product of its mass and the gravitational force acting on the body

i.e. weight =mass *gravitational force.

3.2.2: Tension

When a string is stretched or compressed at both of its ends, it experiences a force called tension. A stretched or compressed material will tend to regain its original shape when the stretching or compressing force is withdrawn. Materials that do not break after stretching or compression is said to be elastic.

3.2.3: Upthrust

When an object is immersed in a fluid (liquid/gas), there is always an opposite upward force acting on it. This is called upthrust force.

NB. There is no upthrust force in a vacuum.

Upthrust force = Weight of object when in air -- Weight of the object when immersed in fluid (Apparent weight of the object)

3.2.4: Friction force

This is the force that tends to oppose the motion of one object over another when they are in contact. Friction force is useful in many ways for instance during walking, writing, applying brakes, lighting a match stick, etc. Friction in fluids is called viscous drag or simply viscosity.

3.2.5: Magnetic force

A magnet can either attract or repel a magnetic material. This force of attraction or repulsion is called magnetic force. Materials that are not affected by a magnet are said to be non-magnetic materials.

3.2.6: Electrostatic force

This is the force of attraction or repulsion between static charges. Like charges repel while unlike charges attract. Electrostatic force is evident in the following cases:

A plastic pen or ruler rubbed on dry air picks up small pieces of paper placed on a table.

Such a pen can attract a stream of water from a water tap.

A glass window wiped using a dry piece of cloth on a dry day immediately attracts dust particles.

Brushed shoes attract dust particles.

When combing a dry hair, a cracking sound is produced.

All these are as a result of attraction between the formed charges and the opposite charges.

3.2.7: Centripetal force

This is the force that constraints a body to maintain a circular path as it move. It is usually directed towards the centre of the circle.

3.2.8: Action and reaction

Action and reaction are equal and opposite forces. When a block of wood rests on a table its weight is exerted on the table. This is action force. The table on the other hand exerts an equal upward force on the block of wood. This is reaction force and it prevents the block from sinking down below the table.

3.2.9: Cohesive and adhesive force

Cohesive force is the force of attraction between molecules of the same kind e.g. force between water molecules. Adhesive force is the force of attraction between molecules of different kinds e.g. force between water molecule and the surface of glass.

For instance when a molecule of water is put on a clean glass surface, the water spreads on (wets) the glass surface. This is because there is greater adhesive force between water and glass molecule than the cohesive force between the water molecules. However, when the glass surface is waxed, the water forms spherical balls. This is because for a waxed surface, the adhesive force between water and glass is lower than the cohesive force between water molecules.

Mercury on the other hand will form spherical balls both on a clean and waxed glass surface. This is because cohesive force between mercury molecules is greater than adhesive force between mercury and glass molecules. Waxing the glass surface in this case even lowers the adhesive force further.

Cohesive and adhesive force can be used to explain the meniscus of water and mercury in a glass tube.

Fig. (i)

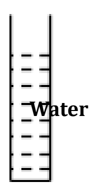
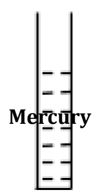


Fig. (ii)



In fig (i), greater adhesive force between glass and water makes water to rise up the narrow glass tube by capillary action so that as many water molecules as possible can be in contact with glass. Hence the meniscus of water curves downwards. The narrower the glass tube the higher the level of water.

In fig (ii), cohesive force between the mercury molecules is greater than the adhesive force between mercury and glass. Mercury thus sinks down the tube so that its molecules can stick together. Hence the meniscus of mercury curves upwards from the glass surface. The narrower the tube, the lower the level of mercury in the tube.

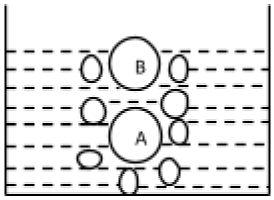
3.2.10: Surface tension

The cohesive force between the molecules of a liquid gives it some strength. The surface can thus resist stretching. Surface tension is the force that makes the surface of the liquid to behave like a fully stretched elastic skin.

Surface tension allows light insects to rest over water without sinking. a steel razor blade or needle floats on water if placed carefully but sinks when soap solution or kerosene is placed around it. Also if one end of the needle is pressed into the water, it breaks the surface tension and the needle sinks.

Molecular explanation of surface tension

Consider two molecules, A and B shown below:



A molecule A deep in the water is surrounded by other molecules from all sides. The net force on this molecule is thus zero i.e. the forces cancel out. A molecule B at the surface has fewer molecules on the upper part. The inward resultant force on B thus causes the surface of water to be under tension.

Note:

Different liquids have different strengths of surface tension.

Activity 1

1. Describe the behavior of soap bubble when blown to the wider end of a funnel.
2. Explain why a glass tumbler can be filled with water above the brim without pouring out.
3. Explain why brush bristles would spread when it is dipped in water but cling together when taken out of water.

3.3: Factors affecting surface tension

3.3.1: Temperature

When the temperature of a liquid is increased the kinetic energy of its molecules is also increased. The inter-molecular distance increases lowering the cohesive force between the liquid molecules. Consequently the surface tension of the liquid is reduced.

3.3.2: Impurities

The presence of impurities in a liquid lowers its surface tension. Examples of impurities include kerosene and detergents like soap solution.

3.4: Mass and weight

Mass is defined as the quantity of matter in a substance. Weight on the other hand is the pull of gravity on an object. Weight of a body depends on the mass of the body and the gravitational field strength at the place. The earth is flatter at the poles such that the distance between the centre of the earth and the poles is shorter than the radius of the earth at the equator. This implies that the force of attraction between the surface of the earth and its centre is greater at the poles than at the equator. Hence a body will weigh more at the poles than at the equator.

The table below summarizes the differences between mass and weight:

Mass	**Weight**
Is the quantity of matter in a body	Is the pull of gravity on a body
Measured in kilogram	Measured in newton
Constant/same everywhere	Varies from place to place

Measured by a beam balance	Measured by spring balance
Scalar quantity	Vector quantity

A scalar quantity is a quantity that can be described in terms of magnitude (size) only eg speed, distance, area etc while a vector quantity is a quantity that has both magnitude and direction e.g. displacement, velocity, momentum etc.

Example 3.1

1. A body weighs 75N on the earth's surface. Calculate its mass in grams if g= 10 N/kg.

W=mxg

m=W/g= 75N/10Nkg^{-1}

=7.5kg=7500g

2. A man weighs 900N on earth and 150N on the moon. If the earth's gravitational field strength is 10N/kg, determine the moon's gravitational field strength.

Mass on earth= W/g=900N/10Nkg^{-1}

=90kg=mass on the moon

Gravitational field strength on the moon=W/m=150N/90kg

=1.67N/kg

3.5: Measurement of force

The most common instrument for measuring force is the spring balance. It uses the principle of extension when there is a stretching force. Some spring balances are calibrated in newtons (N) while others are calibrated in kilogram or even gram. In the latter case, it is advisable to convert the kg or g into newtons.

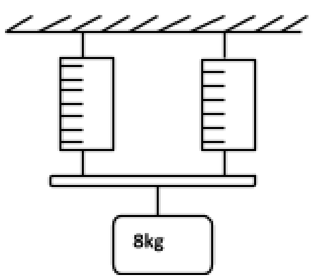
Two or more spring balances can be assembled to act as one spring balance. When the springs are joined in series, the combined extension will be the sum of the individual spring extension.

However, when they are connected in parallel, the springs will share the load. The combined extension therefore will be the extension of one spring divided by the number of such springs in parallel.

Example 3.2

1. The figure below shows two identical spring balances supported as shown:

A **B**

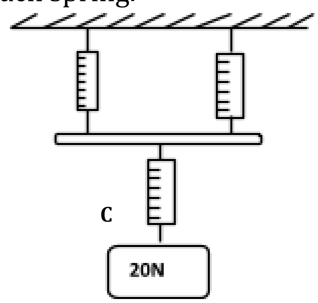


State the reading on each spring balance.

Each spring will read =80/2=40N

2. Three identical arranged as shown below were used to support a load of weight 20N. If the beam has a weight of 1N and each spring would extend by 1cm if a load of weight 4N is suspended from it, determine the extension of each spring.

A **B**



Extension in spring A= Extension in spring B={(21/2)x1cm}/4N =2.265cm

Extension in spring C= (20Nx1cm)/4N =5cm

N/B. Add scalar and vector quantities as well as Problems involving W:MxG

Scalar quantity- a quantity with magnitude only e.gvolume,area,temperature,luminous intensity

Vector quantity-a quantity with both magnitude and direction e.gforce,displacement,velocityetc

TOPIC 4: PRESSURE

4.1: Introduction

Pressure is defined as the compressive force acting normally (perpendicularly) per unit area;

i.e. **pressure= force/area**

The SI unit of pressure is newton per square metre (N/m^2). Pressure can also be expressed in pascals (Pa);

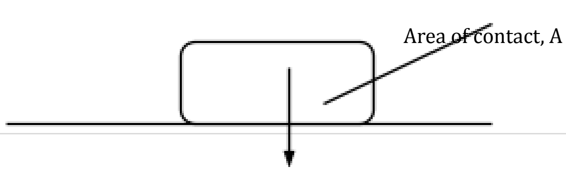
$1N/m^2$=1Pa

Atmospheric pressure is sometimes expressed as mmHg, cmHg or atmospheres.

For a given amount of force, the smaller the area of contact the greater the pressure exerted. This explains why it would be easier for a sharp pin to penetrate a piece of cardboard than a blunt one when the same force is used.

4.2: Pressure in solids

A solid resting on a horizontal surface exerts a normal contact force equals to its weight. The pressure of the solid on the surface depends on the area of contact.



Weight, W

Example 4.1

1. A man whose mass is 90kg stands on a floor.

   a) If the area of contact between his feet and the floor is $0.0368m^2$, determine how much pressure he able to exert on the floor.

   $P= F/A = 900N/0.0368m^2$

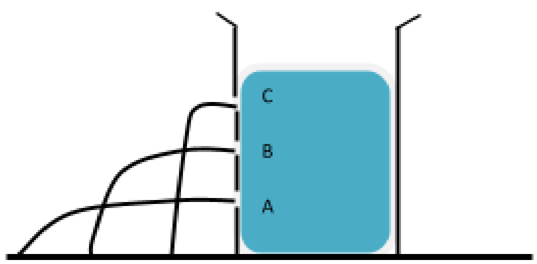
   $=24,456.5217N/m^2$.

   b) What pressure will he exert on the floor if now he stands on one foot?

   $P=900N/(0.0368/2) =48,913.0435N/m^2$

4.3: Pressure in fluids

A fluid is a term that refers to either liquids and gases. The set up below can be used to illustrate pressure in fluids:

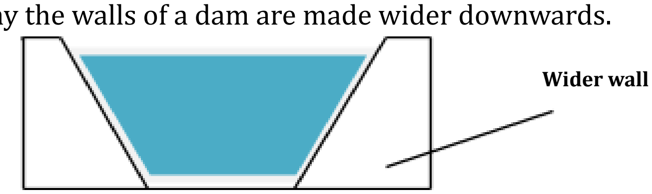


The lower hole **A** is observed to throw water the farthest, followed by hole **B** and **C** the closest. This indicates that pressure at **A** is greater than that at **B** and pressure at B is greater than that at **C**.

Conclusion

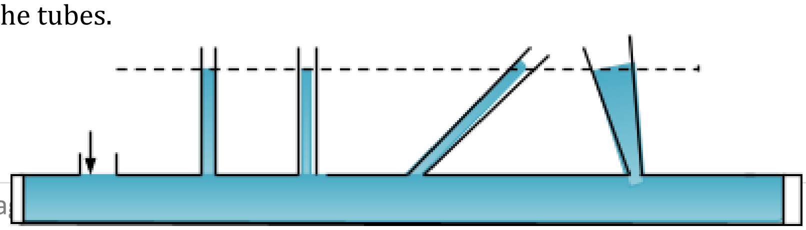
Pressure in fluids increases with depth i.e. the greater the depth the higher the pressure it exerts.

This explains why the walls of a dam are made wider downwards.



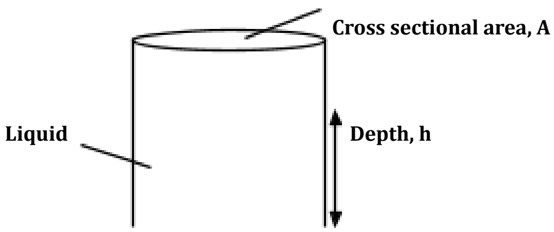
A diver under water experiences pressure due to the weight of water above him plus the atmospheric pressure above the water surface. The deeper the diver, the greater the pressure.

When a liquid is poured into a set of connected tubes of different shapes, it rises up until the levels are the same in all the tubes.



4.4: The fluid pressure formula

Consider a liquid of density ρ in a container of uniform cross-section area A, such that the depth of the liquid in the container is h;



Volume of the liquid = A*h

Mass = volume * density

=Ahρ

Weight of the liquid = mass * gravitational field intensity= force exerted

=Ahρg

From the definition of pressure, P= F/A

= Ahρg/A

= hρg

It is thus clear that pressure in fluids is directly proportional to the height of the column h, the density of the fluid ρ and the gravitational field strength g.

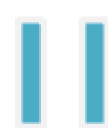
Note: Pressure in fluids does not depend on the cross section area of the container which holds it.

Example 4.2

1. Calculate the pressure exerted by a column of kerosene of 850mm. take the density of kerosene= 800kgm^{-3}.

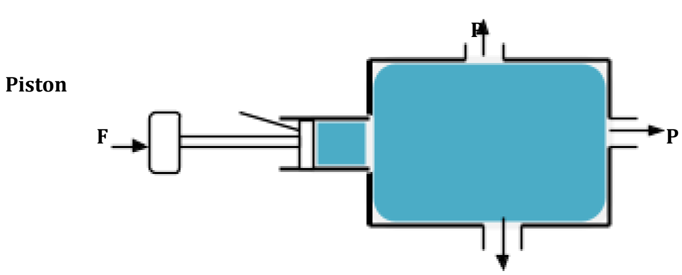
P=hρg =0.85m x 800kgm^{-3}x10N/kg

= 6800Pa



4.5: Transmission of pressure in fluids

The figure below shows a liquid under pressure due to the force F acting on the plunger.



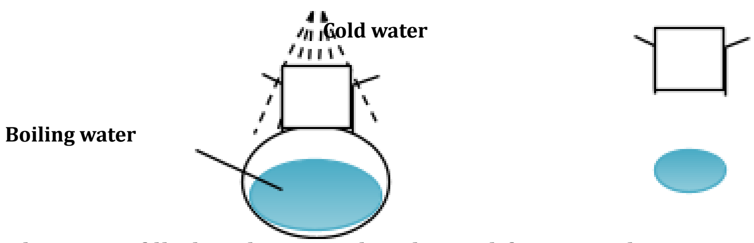
P

Assuming that the holes are identical, when the plunger is pushed forward, the liquid squirts out through the holes with equal force. If the piston area is **A**, then the pressure developed is **F/A**. this pressure is transmitted equally to all parts of the liquid. This is called Pascal's principle. The principle states that pressure applied at one part of a liquid is transmitted equally to all other parts of the enclosed liquid. Gases can also transmit pressure in a similar way provided they are incompressible.

The working of hydraulic machines is based on Pascal's principle.

4.6: Atmospheric pressure

The term atmosphere refers to the air surrounding the earth. The weight of air above the earth's surface exerts pressure on the earth. This pressure is called atmospheric pressure. The presence of atmospheric pressure can be demonstrated by the crushing can experiment;



The can is filled with water then heated for several minutes. After sometime, the can is sealed and then cooled by running cold water over it.

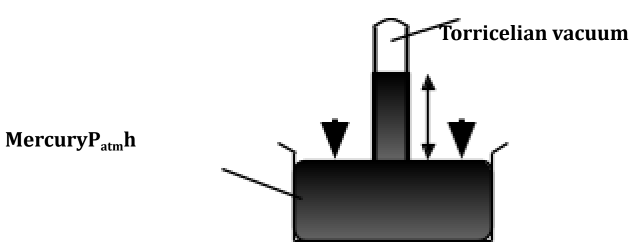
When the water is heated, steam is produced which displaces air in the can. When cold water is run over it, steam condenses leaving a vacuum in the can. Pressure inside is thus reduced below the external atmospheric pressure. Hence the can crushes inwards.

Atmospheric pressure is also very vital when using a drinking straw. By sucking through a drinking straw the pressure inside is reduced. The atmospheric pressure acting on the surface of the liquid overcomes the pressure inside the straw. The pressure difference and hence the resultant force pushes the liquid up the straw.

4.7:Measurement of pressure

Atmospheric pressure is measured using an instrument called a barometer. The following are some types of barometers:

- **A mercury barometer**



Atmospheric pressure can support a liquid column in a tube. One end of the tube is closed and the tube is filled with mercury. When inverted and with the open end below the liquid surface in the container. The atmospheric pressure (P_{atm}) on the open surface is transmitted by the liquid to the base of the liquid column and supports its weight.

With the liquid column h, the atmospheric pressure can be determined from the equation;

Pressure = hρg

Where h- is the height of the liquid column

ρ- Density of the liquid (mercury)

g- Gravitational field strength

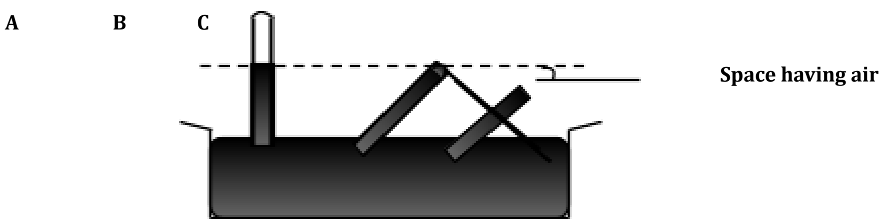
At sea level atmospheric pressure can support approximately 76cm column of mercury equivalent to approximately 10m column of water. Mercury is thus preferred as a barometric fluid since it gives a shorter and measurable column compared to water.

In general, atmospheric pressure decreases with altitude. The value of atmospheric pressure at sea level is called the standard atmospheric pressure and is at times referred to as one atmosphere.

Pressure at sea level =hρg =0.76m*13600kg/m^3 *10N/kg

= 103,360N/m^2

If there is air trapped in the space above the mercury column then the barometer is faulty. This space above the mercury column is called toricellian vacuum. To test whether this space has some air trapped, the test tube is tilted until it is at the same level with the mercury column when the tube is upright. If the space is truly a vacuum, the test tube will be completely filled with mercury while if it has trapped air a space will still remain at the top.

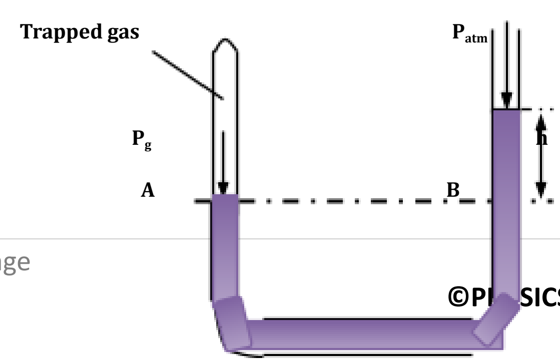


Normally the toricellian vacuum contains some little mercury vapour.

Note that this barometer is not readily portable.

- **A manometer**

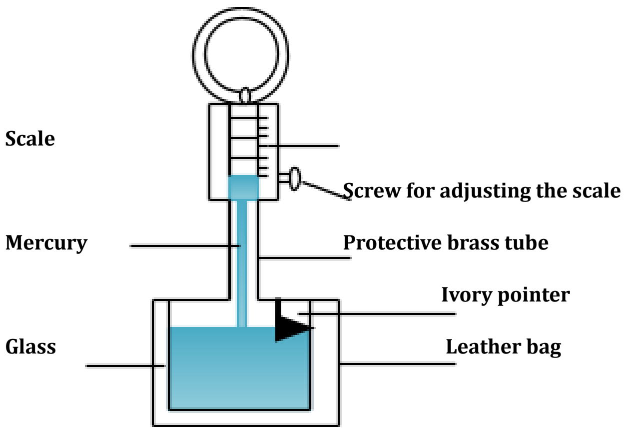
This is a U-shaped tube open on both ends. One end is connected to a source of gas whose pressure is to be determined. The other arm is open to the atmosphere. This creates a pressure difference which displaces the manometer liquid.



The points A and B are at the same level and as such experience the same amount of pressure. The pressure at A is the gas pressure while that at B equals the pressure due to the liquid column plus atmospheric pressure;

$P_g = P_A + h\rho g$

- **Fortin barometer**

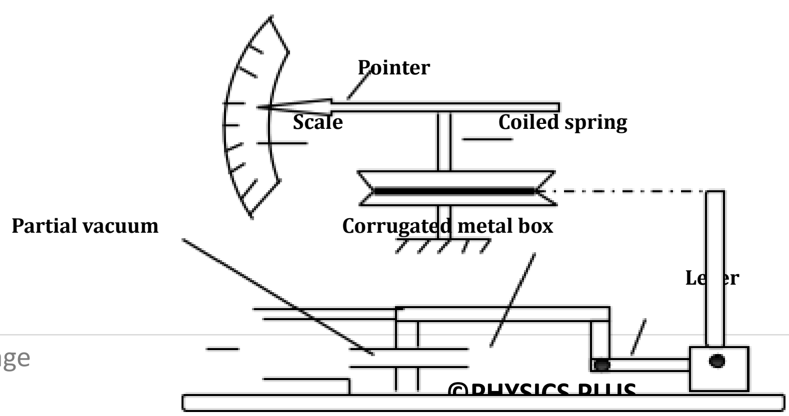


The fixed ivory index with a sharp point acts asthe zero mark of the main scale. Before taking any reading the level of the mercury in the reservoir must first be adjusted until the tip of the ivory index just touches the surface of mercury.

The height of mercury column is then read from the main scale .this reading is then used to calculate the pressure at the place. Any change in the atmospheric pressure causes the level of mercury in the reservoir to move up or down, hence the adjustment of the ivory index is necessary.

- **Aneroid barometer**

This type of barometer is more portable.



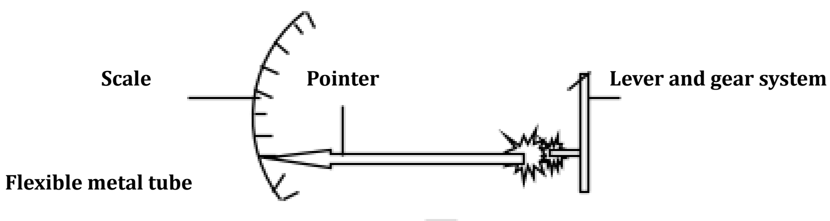
Strong spring

When the pressure outside the corrugated metal box is reduced, the box expands setting the levers into motion. However, when the pressure outside increases, the box reduces in volume. The resultant movements of the springs and levers moves the pointer across the scale recording the value of the atmospheric pressure.

The aneroid barometer can also be used to measure heights. For instance, altimeters are aneroid barometers used in aircrafts to measure heights.

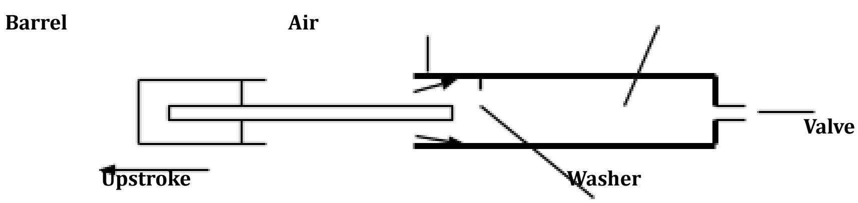
- **Pressure gauge**

Pressure gauges are also easily portable. It is commonly used to measure gas pressure, tyre pressure, etc. It consists of a coiled flexible metal tube. When the pressure inside the tube increases, the tube uncoils. The movement of the tube is magnified by the lever and gear mechanism which then moves the pointer across the scale.



4.8: Applications of pressure

1. **A bicycle pump**



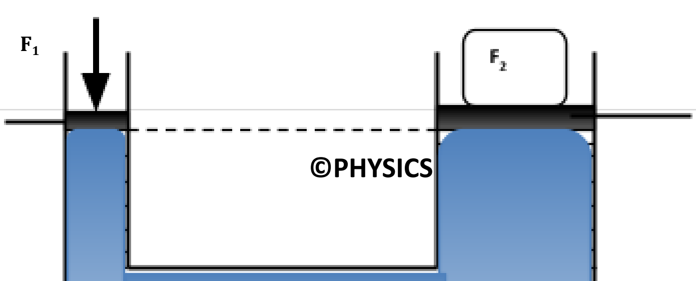
The leather washer is very flexible and works both as a valve and a piston. During the upstroke, air in the barrel expands and its pressure reduced below the atmospheric pressure outside. The pressure difference and hence the resultant force acting downwards pushes the air past into the barrel. The valve remains closed due to the high pressure in the tube.

During down stroke, air in the barrel is compressed raising its pressure. This high pressure presses the leather washer against the wall of the barrel hence no air leaks out. When the pressure of the air in the barrel overcomes that inside the tube, the air is forced through the valve into the tube. The work done in compressing the air in the barrel generates some heat raising the temperature of the barrel.

2. **The hydraulic machines**

These machines apply Pascal's principle in their working. They include the following machines:

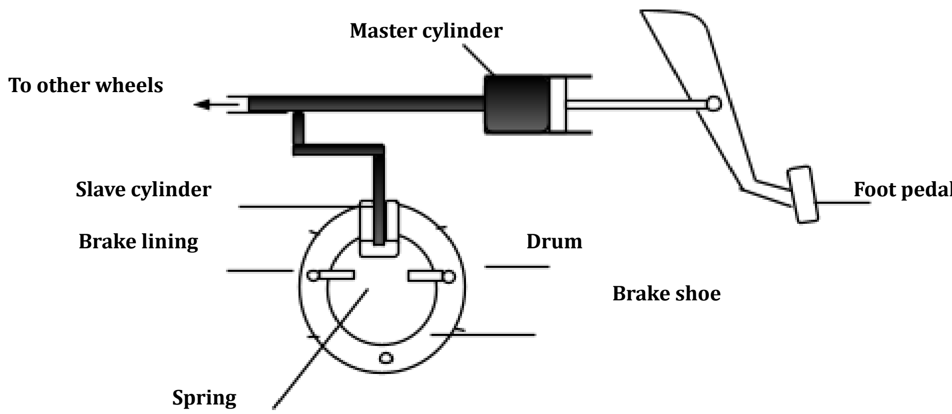
a) Hydraulic lift



Piston (area A_1) **Piston (area A_2)**

When a force F_1 is exerted on the smaller piston of cross section area A_1, the pressure developed (F_1/A_1) is transmitted by the liquid to the larger piston of cross section area A_2 and is able to support a load F_2.Alternatively, the pressure exerted on the liquid by the larger piston can be expressed as F_2/A_2. If the system is in equilibrium, then $F_1/A_1 = F_2/A_2$.

b) Hydraulic brake system



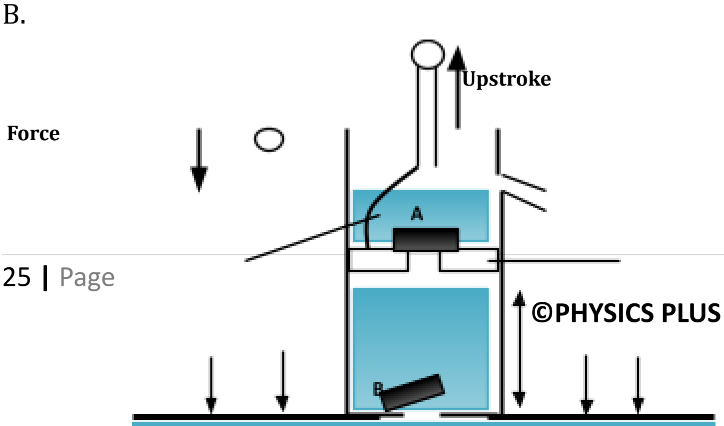
The force applied on the foot pedal exerts pressure on the master cylinder. This pressure is transmitted equally by the brake fluid to the slave cylinder. The pistons of the slave cylinder are then forced to open outwards. This opens the brake shoe. The brake lining then presses the drum, resisting the rotation of the wheel. When the force on the foot pedal is released, the return spring pulls back the brake shoe and the wheel can rotate once again .The pressure exerted on the master cylinder is transmitted equally to all the four wheels and so the braking force is uniform

In order for the brake system not to fail, the liquid used should have the following properties;

- ✔ Be incompressible.
- ✔ Have low freezing point and high boiling point.
- ✔ Should not corrode parts of the brake system.

3. The lift pump

The lift pump relies on the atmospheric pressure to raise water from a borehole or well. It has two valves, A and B.



Piston **Plunger**

P_{atm} **hP_{atm}**

To start the pump, water must be poured on top of the piston, a process called priming. This creates an air-tight seal around the piston and in valve A. The pump is operated by means of a lever. It has two cycles, upstroke and down stroke.

During upstroke

When the plunger moves up during upstroke, valve A closes due to the pressure of water above it plus the weight of the plunger. At the same time, air above valve B expands and the atmospheric pressure on the surface of water pushes water past valve B into the barrel. The plunger is moved up and down until the region between A and B is filled with water. Water above the piston is then lifted out through the side tube.

During down stroke

During downstroke, valve B closes due to its weight plus pressure of water above the piston.

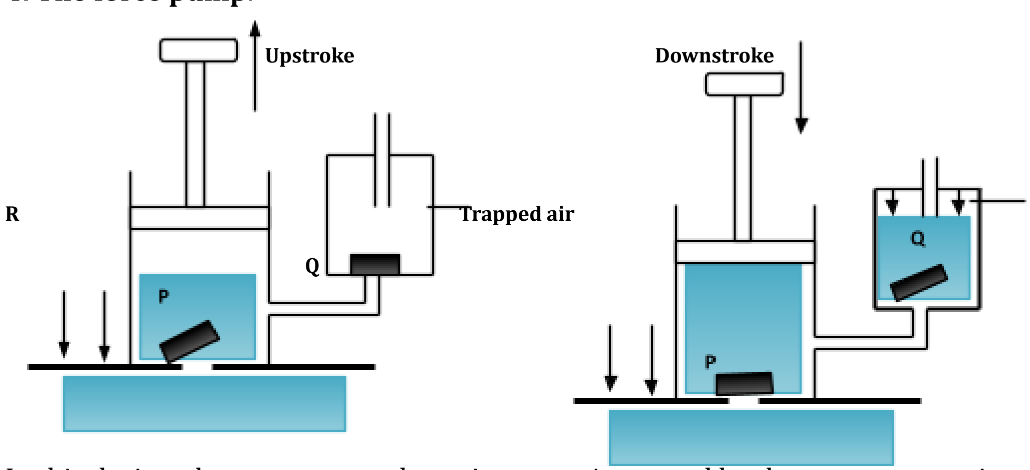
Limitation of the lift pump.

The atmospheric pressure can only support a column of water of about 10mm i.e. Normal atmospheric can raise water to a maximum height of 10m.

In practice, this height is usually below 10m.This due to;

- ✔ Low atmospheric pressure in places at high altitudes.
- ✔ Leakages at the valves and piston

4. The force pump.



In this design, the pressure used to raise water is created by the person operating the pump handle. The pump handle moves a piston in the cylinder, which is placed within 10m of the water level i.e. 10m or below [h].The cylinder is connected to a chamber.

Upstroke.

During upstroke, air above valve P expands and its pressure falls below atmospheric pressure. The atmospheric pressure on the water surface pushes water up past valve P into the barrel.
The pressure above valve Q is atmospheric pressure and hence this valve remains closed during this stroke.

Down stroke

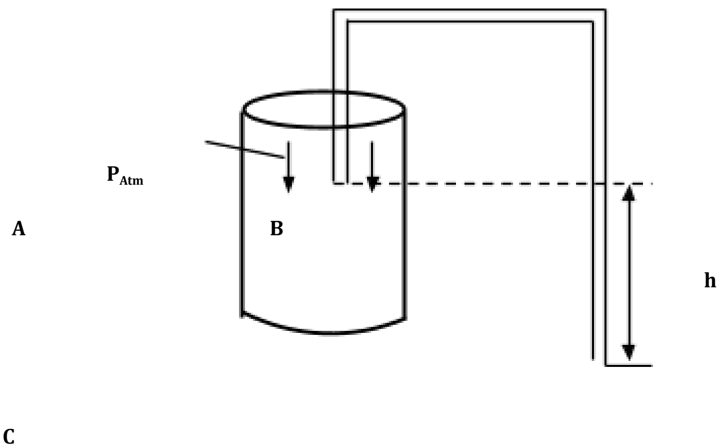
During down stroke, valve P closes. Increase in pressure in water in the barrel opens valve Q and water is forced into the chamber R. As water fills the chamber, some air is trapped and compressed at the upper part.

In the next upstroke, valve Q closes and the compressed air forces water up the delivery pipe ensuring a continuous flow of water as the valves open and close.

NB: The height to which water can be raised is affected by two factors, namely:

- The strength of the person pumping i.e. the force applied during the down stroke
- The ability of the pump and its working parts to withstand the pressure of the long column of water in the chamber R.

5*. Siphon*



The pressure at A and B is the same and equal to the atmospheric pressure. The pressure at C is the atmospheric pressure plus the pressure due to the column of water, h.

$P_c = P_{atm} + h\rho g$

The pressure of the surrounding air at C is atmospheric pressure. This creates a pressure difference. The resultant downward force due to this pressure difference causes water to flow out of the end C. The water column doesn't break due to strong cohesive force between water molecules.

For the siphon to work, the following conditions must be met:

- The end C must be below the surface of the liquid to be emptied, A.
- The tube must first be filled with the liquid having NO bubble in it.
- One end of the tube must be in the liquid
- The tube must not rise beyond the barometric height of the liquid from the surface of the liquid.

Note that a siphon can work in a vacuum provided a pressure difference is created.

TOPIC 5: <u>PARTICULATE NATURE OF MATTER</u>

5.1: Introduction

Matter – is anything that occupies space and has mass. Matter exists in three states namely; solid, liquid and gas.

Experiment

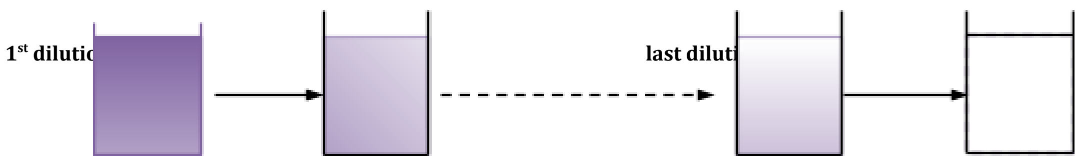
Aim: to demonstrate that matter is made up of smaller particles

Apparatus: Beakers, potassium permanganate crystals, water.

Procedure

i. Pour water into the beaker till it is half full
ii. Dissolve some potassium permanganate crystals until the solution is purple
iii. Transfer half the solution into another beaker and fill it with water
iv. Continue the process with other beakers, comparing the colour of the solution through each dilution.

Observation



The colour of the solution fades away through each dilution until the solution appears colourless.

This shows that the particles of potassium permanganate are spread out evenly in the water. Each dilution spreads them out further as the water molecules increase in number. Hence the purple colour fades away with each dilution until it becomes colourless.

Conclusion

Potassium permanganate is made up of tiny particles.

Also, when some salt particles are put in a flask and water added, it is observed that after shaking the flask to dissolve the salt, the volume of the final solution is less than the original volume of the water plus undissolved salt. This is because salt particles fitted into some spaces between the water molecules. Hence the particles of the solution are more packed together, reducing the volume of the solution. This also indicates that the particles of water and those of salt differ in size.

5.2: Brownian motion

Brownian motion refers to the irregular movement of light particles when they are knocked by heavier bodies.

Experiment 1

Aim: To demonstrate Brownian motion in liquids

Apparatus: Beaker, hand lens, pollen grains (chalk dust), transparent lid.

Procedure

i. Pourwater into the beaker about ¾ full.

ii. Sprinkle pollen grains on the surface of the water.

iii. Cover the beaker with a transparent lid.

iv. By the use of a hand lens, observe the behavior of the pollen grains on the water surface.

Observation

The pollen grains suspended in the water are in constant random motion.

The pollen grains are constantly knocked by smaller invisible particles of water from all sides. The net force on the pollen grain at any instant makes it to move with an irregular pattern. This is called Brownian motion.

Experiment2

Aim: to demonstrate Brownian motion in air (the smoke cell experiment)

Apparatus: A drinking straw**,** smoke cell, microscope and a bright light source

Procedure

i. Burn one end of the straw and let the smoke fill the smoke cell from the other end.

ii. Remove the straw and cover the cell using the cover plate.

iii. Illuminate the cell and with the use of a microscope observe the behavior of the smoke particles in the smoke cell.

Observation

Bright specks which are in continuous random motion are observed.

The bright specks are smoke particles scattering light. The random motion is due to constant collision between smoke particles and invisible air particles and also with walls of the cell.

The above findings can be summarized by the **kinetic theory of matter** which states: **matter is made up of tiny particles which are in continuous random motion**.

5.3: States of matter

Matter exists in three states which are interconvertible through either heating or cooling.

5.3.1: Solids

In solid state, the particles are closely packed together due to strong cohesive force between them. They vibrate in their fixed positions. Hence solids have definite shape and volume. When the temperature of the solid is raised, the vibration becomes rapid. At a certain temperature the particles of the solid break away from the structure of the solid and the solid is said to have melted. This temperature is called the **melting point** of the solid.

5.3.2: Liquids

The particles of a liquid are generally not as close as in solids. The attractive force between the liquid particles is weaker than solids'. Hence liquids have neither definite shape nor volume but take the shape of the container.

The density of a liquid is a little less than that of its liquid. This is because the liquid occupies more space than its own solid. Hence solid materials sink in their own liquids.

When a liquid is heated to a certain temperature, it changes to a gas. This temperature is known as its **boiling point** and the process is called **vaporization**.

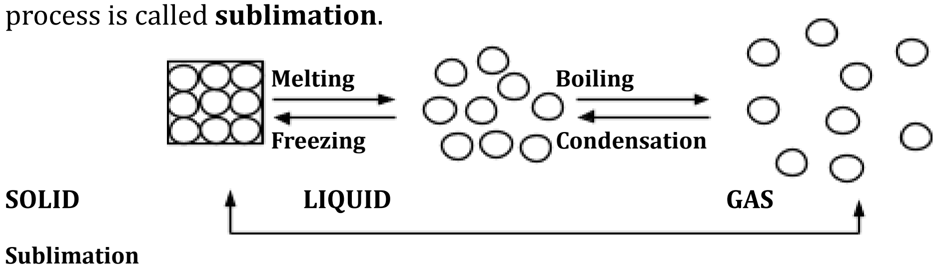
A liquid can sometimes lose its energy and fall back to solid state. This is called **freezing**.

5.3.3: Gases

In gases, the particles are far apart with the weakest cohesive force. For this reason gases are highly compressible. Gases have densities which are much less than those of their respective liquids and solids.

Gas particles can sometimes lose their energy and fall back to liquid state. This is called **condensation**.

NB: certain solids change directly to gas when heated while some gases can be cooled directly to solid. This process is called **sublimation**.



5.4: Diffusion

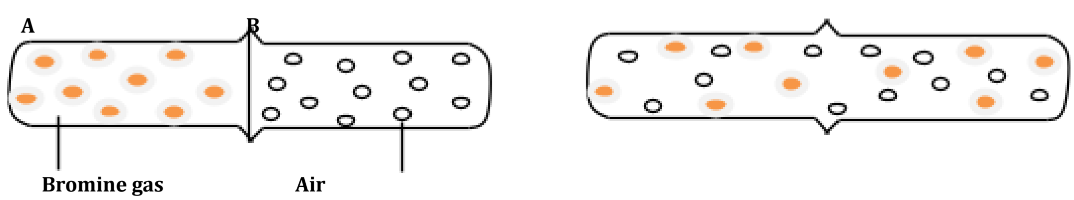
Diffusion is defined as the process by which particles move from a region of high concentration to a region of low concentration.

- Diffusion in liquids can be demonstrated by putting a potassium permanganate crystal at the bottom of a beaker of water. After some time the water is observed to turn deep purple at the bottom and light purple near the top.

Also when a saturated copper II sulphate solution is added to water in a beaker, initially the water layer floats on top of the copper II sulphate solution due to its lower density. After some time the boundary disappears and the two form a homogeneous pale blue mixture.

When warm liquids are used in the illustrations above, similar observations would be made but after a shorter time. This suggests that temperature speeds up the rate of diffusion.

- Diffusion in gases is faster than in liquids due to their lower density, high kinetic energy and weaker cohesive forces (or larger intermolecular distance). Diffusion in gases can be investigated by setting up two gas jars, one filled with air and the other with bromine gas as shown below.



After some time a pale brown gas mixture is observed in both the jars. The bromine gas spreads into jar B while air particles also spread into jar A.

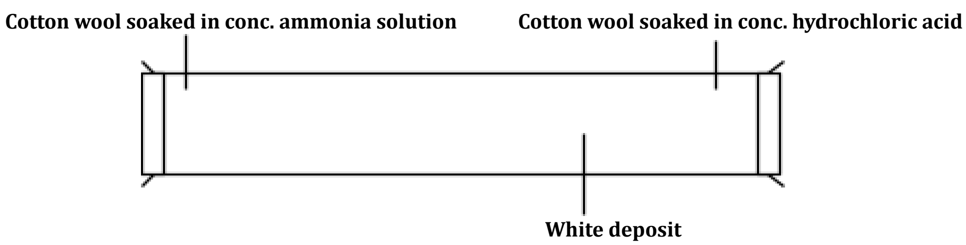
NB: The rate of diffusion will be slower when the gas jars were set upright. This is because the denser gas will tend to remain at the base and the lighter one up.

This experiment should be carried out in a fumed cupboard or fume chamber.

- Diffusion in solids is very slow but can occur between two metals like gold and lead.

5.4.1: Rate of diffusion

The rate of diffusion between ammonia gas and hydrochloric acid gas can be investigated by the set up below.

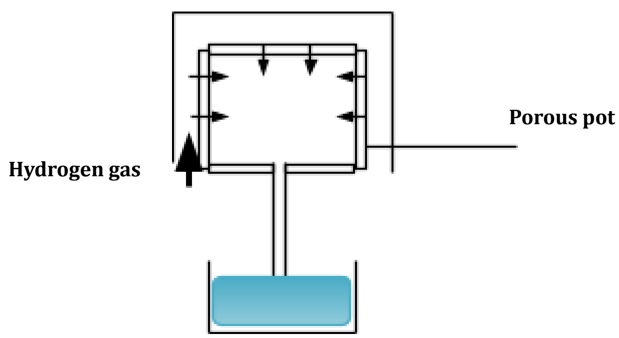


After some time, a white deposit of ammonium chloride is formed near the end with the hydrochloric acid. This indicates that ammonia gas diffuses faster than hydrochloric acid gas. Note that ammonia gas is lighter compared to hydrochloric acid gas.

Conclusion

The rate of diffusion depends on the **density** of the substance and the **temperature**.

5.4.2: Diffusion through porous materials



A porous pot has very fine holes through which hydrogen gas diffuses into the pot while air diffuses out. The level of water in the tube falls as hydrogen diffuses into the pot. If hydrogen is supplied for a longer time, bubbles of hydrogen gas will be observed from the end in the water.

When the gas supply is stopped, hydrogen gas diffuses out of the pot as air diffuses into the pot and the level of water in the tube rises.

TOPIC 6: <u>THERMAL EXPANSION</u>

6.1: Introduction

Temperature – it is the degree of coldness or hotness of a body on some chosen scale. It is measured using a thermometer. The SI unit of temperature is the Kelvin (K). Other units include degrees Celsius (^{0}C), Fahrenheit, F. Temperature is a basic physical quantity as well as a scalar quanity.

6.2: Temperature scale

The scale of a thermometer is obtained by selecting two temperatures called fixed points; the lower fixed point and the upper fixed point. The lower fixed point is the temperature of pure melting ice. It is taken to be 0^{0}C. The upper fixed point is the temperature of steam above pure boiling water at normal atmospheric pressure. It is taken to be 100^{0}C. The temperature of steam is used since impurities do not affect its temperature but will raise the boiling point of water. The range between these two points is then divided into equal divisions.

On the Kelvin (absolute) scale, 0^{0}C is at 273 K while 100^{0}C is at 373 K. Hence to convert ^{0}C to K, add 273 to the temperature in ^{0}C.

Activity 6.1

1. Convert the following into Kelvin:

a) 35^{0}C b) -111^{0}C c) -273 ^{0}C

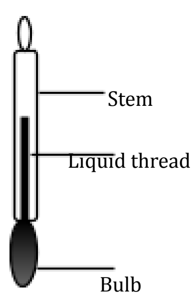
2. Convert the following into ^{0}C:

a) 123 K b) 323 K

6.3: Types of thermometers

A thermometer is designed according to the purpose for which it is required. The following are some of the commonly used thermometers:

6.3.1: Liquid-in-glass thermometer



In this thermometer the liquid expands up a capillary tube when the bulb is heated. The liquid used in this thermometer should posses the following qualities for the thermometer to be effective:

- ✔ Be easily visible
- ✔ Expand and contract uniformly
- ✔ Have a wide range of temperature i.e high boiling point and low freezing point
- ✔ Be sensitive to small temperature changes
- ✔ Should not wet the glass

The most commonly used liquid is mercury althoughcoloured alcohol can also be used. Water does not meet all the above desirable properties. The table below compares mercury and alcohol as a thermometric liquid:

Mercury	**Alcohol**
Has high b.p, 357^{0}C	Has low b.p, 78^{0}C
Relatively high freezing point, -39^{0}C	Low freezing point, -115^{0}C
Good thermal conductor	Poor thermal conductor
Has regular expansion	Has a slight irregular expansion
Does not wet the glass	Wets the glass
Easily visible	Ii coloured to make it easily visible

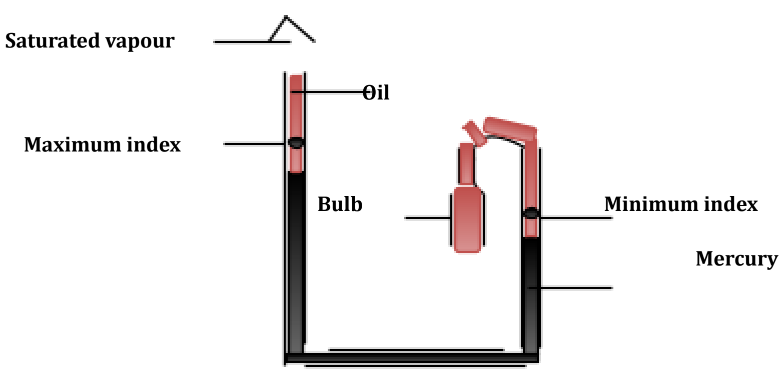
These thermometers are commonly used in normal laboratories.

6.3.2: A clinical thermometer

This is a special thermometer used to measure human body temperature. It has a short scale between 35—43^{0}C. This is because the optimal body temperature is 37^{0}C. It has a constriction to prevent back flow of the mercury into the bulb. This is to allow time to take the reading. After use the thermometer is shaken to return the mercury back to the bulb.

Methylated spirit can be used to sterilize the thermometer after use.

6.3.3: Six's maximum and minimum thermometer



It is used to record minimum and maximum temperatures of a given place. When the temperature of the surrounding rises, the oil in the bulb A expands pushing the mercury which in turn pushes up the oil in the other arm. This compresses the vapour above the oil and the maximum index is pushed up to the maximum position. This is the maximum temperature.

When the temperature falls, the oil contracts back into the bulb. Mercury flows back pushing the minimum index to the minimum position. This gives the minimum temperature.

After taking the readings, the indices are pulled down to the level of the mercury using a magnet.

6.3.4: A bimetallic thermometer

It is made up of a coiled bimetallic strip whose one end is fixed and the other end connected to a pointer. Commonly used metals are brass and invar. When the temperature rises brass expands more than invar. The strip thus curls forcing the pointer to move over a calibrated scale.

6.4: Expansion in solids

A solid can expand in three ways:

- ❖ Linear expansion; increase in length
- ❖ Superficial expansion; increase in surface area
- ❖ Cubic expansion; increase in volume

Solids expand when heated and contract when cooled. During expansion the volume increases, density decreases but mass remains the same. Expansion in solids can be demonstrated by the ball and ring experiment.

When both the ball and ring are at room temperature, the ball easily passes through the ring but when the ball is heated it does not go through the ring. When left in contact for some time the ball finally passes through the ring again.

On heating the ball expanded and so could not go through the ring. After sometime it went through because the ball lost some of its heat to the ring which then expanded while the ball slightly contracted.

Different solids like metals will expand at different rates when exposed to the same amount of heat for the same duration. This can be investigated by the bar and gauge experiment.

One end of the metal bar is fixed while the other end is kept in contact with the pointer. Any slight expansion of the bar is magnified by the long pointer and can be read from the scale. The experiment is then repeated using bars of other materials. The pointer readings are then used to compare their rates of expansion.

In the above experiment, the following parameters must be kept constant:

- ✔ Length of the rods
- ✔ Diameter / thickness of the rods
- ✔ Source of heat
- ✔ Duration of heating

The measure of the tendency of a material to expand is called its expansivity. The ability of a material to expand when heated is referred to as its linear expansivity.

Linear expansivity, α = expansion (change in length)/ {original length * temperature change}

The SI unit of linear expansivity is per Kelvin (K^{-1}).

Linear expansivity of a substance may also be defined as the fraction of its original length by which a rod of the same substance expands per Kelvin rise in temperature.

Example 6.1

1. Consider a brass rod of length 50.2 cm at 16.6^{0}C. if the rod is heated until a temperature of 99.5^{0}C where its new length is 50.279 cm, determine the linear expansivity of brass.

Linear expansivity, $\alpha = e/l_0 * \Delta T$ = (50.279-50.20) cm/50.2 cm * (99.5-16.6) K

= 0.079/50.2-82.9

= $1.9 * 10^{-5} K^{-1}$

The table below shows some substances with their linear expansivities:

Material	**Linear expansitivty ($* 10^{-5}$) K^{-1}**
Aluminum	2.6
Copper	1.68
Brass	1.9
Iron	1.2
Steel	1.1
Concrete	1.1
Platinum alloy	0.9
Glass	0.85
Invar	0.1
silica	0.042

The knowledge of linear expansivity is used in designing various materials to ensure that they are able to operate well under varying thermal conditions. For instance ordinary glass has a higher linear expansivity than a pyrex glass. When hot water is put in an ordinary glass, it breaks but when a pyrex glass is used it does not crack. The pyrex glass has lower linear expansivity and cannot suffer very large forces of expansion while the ordinary glass does as it undergoes temperature changes.

In building and construction, concrete is always reinforced using steel because both have the same linear expansivity.

6.5: Bimetallic strip

It is formed when two metals of different linear expansivitiesare riveted together e.g. brass and iron or brass and invar. When the temperature of the strip is raised, brass expands more than iron. Hence the strip curves with brass curving outwards and iron inwards. When the temperature falls, brass again contracts more than iron and the strip curves with brass now on the inner side and brass on the outer side.

6.6: Applications of expansion and contraction in solids

6.6.1: Railway lines

Railway lines are fixed with gaps to allow for expansion when temperature rises. The bolt holes are also oval in shape for the same reason. Another way of creating room for expansion in railway lines is by planing the ends of the rails so that they are able to overlap during expansion.

6.6.2: Telephone/electricity wires

Telephone and electricity wires are loosely fixed during installation to allow for contraction during cold weather.

6.6.3: Steam pipes

Pipes carrying steam from boilers are fitted with expansion loops to allow for expansion and contraction. Without the loop the pipe is likely to break due to the resultant force as a result of expansion and contraction. It is necessary that oil companies make this allowance when constructing fuel pipelines.

6.6.4: Steel bridges

In the construction of steel bridges, one end is fixed while the other end is placed on rollers. This is to allow for expansion and contraction.

6.6.5: Rivets

Rivets are fitted when hot and then hammered flat. On cooling, the rivet contract, pulling the two plates firmly together.

6.6.6: Thermostat

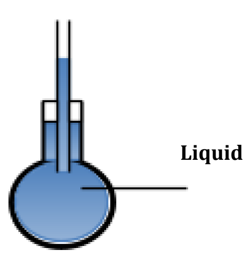
It is a device that can be used to control the temperature of a room. It uses a bimetallic strip. It is connected to a heater circuit. When the temperature of the room rises beyond the set value, the bimetallic strip expands and bends away breaking the contact. Hence the heater circuit is switched off.

The strip cools and contracts and the contact is remade switching on the heater circuit. The setting knob is used to adjust the temperature at which the thermostat is switched on and off.

Other uses of the thermostat include controlling the temperature of electric iron, cookers and fridge, fire alarms and car indicators.

6.7:Expansion and contraction in liquids

The rate expansion in liquids is more than in solids because the particles are slightly far apart. When temperature increases, the liquid molecules gain more energy increasing their rate of movement. The weak bonds between these molecules are further weakened. The molecules thus expand and occupy more space. Expansion in liquids can be demonstrated by the set up below:



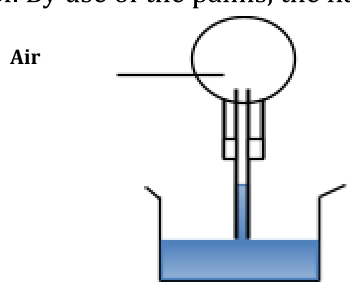
When heated, the level of the liquid in the glass tube first drops and then starts rising. This initial fall in the level is because the glass was heated first and expanded. Later the liquid received the heat energy and expanded hence the rise in the level.

Just like solids, liquids expand at different rates. In order to investigate this, a number of identical flasks are filled with different liquids ensuring that their initial levels are the same in the glass tubes. For a fair comparison, the tubes should be identical i.e. of same diameter. The flasks are then simultaneously immersed in a bath of hot water. The bath of water should be stirred continuously to ensure that temperature is uniform.

It will be observed that the level of the liquids in the tubes differ after some time. If water, alcohol and methylated spirit were used, it would be observed that methylated spirit expanded the most, followed by alcohol and water the least.

6.8: Expansion in gases

Gases have the highest rate of expansion because their particles are very far apart and are held by very weak forces. When heated, they gain more energy and move farther apart occupying more space. It can be shown by a round bottomed flask fitted with a glass tube in a tight-fitting cork. The flask is first inverted with the glass tube dipped in water. By use of the palms, the flask is warmed for some time.



It will be observed that the level of water in the tube drops and if warmed for a longer time, bubbles are observed escaping from the end of the tube in water. This shows that air expanded on heating and needed more space, hence the drop in the level of water in the tube and the bubbles.

If the heat is withdrawn, the level of the water rises again in the tube. Expansion and contraction in gases is the basis of the formation of land and sea breezes.

6.9: The Unusual expansion (anomalous) expansion of water

It is normal experience that substances expand on heating and contract on cooling. But for water, this is never to be between the temperatures 0^0C and 4^0C. Water can exist as a solid (ice), liquid (liquid water) and as a gas (steam).

At temperatures below 0^0C, water exists as a solid, occupying a bigger volume. When heated, it expands just like any other solid up to 0^0C. At 0^0C, ice melts at constant temperature. Melting is accompanied by a decrease in volume by about 8%. Beyond 0^0C, water contracts further up to 4^0C. Therefore water has minimum volume at 4^0C and hence maximum density which is slightly higher than 1 g/cm^3.

Above 4^0C, water expands like any other liquid. This behavior of water is described as anomalous, unusual, or irregular.

The variation of volume with temperature and density with temperature when water is heated is illustrated by the graphs below:

6.9.1: Effects of anomalous expansion of water

a. Biological importance

During cold weather, the temperature of lakes and ponds drops and water contracts, becomes denser and sinks. A circulation of water is thus set up until all the water attain maximum density i.e. at 4^0C. If further cooling occurs (below 4^0C), then any water below 40C will stay at the top due to its lower density. At 0^0C, ice forms on top and this acts as an insulator to the layers below. Hence the warmth underneath can sustain aquatic life and thus the aquatic animals and plants can survive there.

b. Icebergs

Ice has a slightly lower density, about 0.92 g/cm^3, than that of water and hence it floats with a small portion above the water surface. The rest and a bigger portion of the ice rests under water. This is called an iceberg. Icebergs pose a great danger to ships as the submerged parts cannot be seen easily by navigators.

c. Weathering of rocks

Water sometimes finds its way into cracks within the rocks. When such water freezes during cold weather, it expands forcing the rock to break into smaller pieces. This is very important for agriculture as soil is formed.

d. Bursting of water pipes

At times the water flowing through a pipe may freeze when it passes through a cold region. The water thus contracts, expanding and this may lead to pipe bursts if expansion allowances were not catered for.

TOPIC 7: HEAT TRANSFER

7.1: Heat and temperature

Heat may be defined as a form of energy that flows from a hot body to a cold body. Its SI unit is the joule (J). When a body loses heat its temperature is lowered while when a body gains heat its temperature rises. If two bodies which are at the same temperature are in contact, there is NO net heat transfer and the bodies are said to be at thermal equilibrium. No instrument can measure directly the amount of heat on a body.

Temperature on the other hand is the measure of the degree of hotness or coldness of a body. Its SI unit is the Kelvin (K). The most commonly used unit is degrees Celsius (^0C). Heat energy can only flow if there is a temperature difference between the two bodies.

7.2: Modes of heat transfer

There are three modes of heat transfer namely conduction, convection and radiation.

7.2.1: Conduction

Conduction is mainly prominent in **solids**. When a solid is heated, say from one end, the other end also becomes hot after some time. The means by which the heat is conducted to the other end can be explained by two theories:

a) Particle/atomic vibration

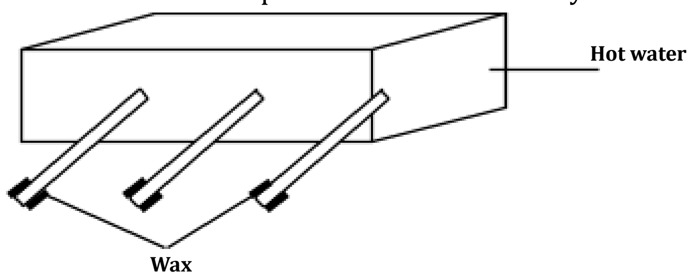
When a material is heated from one end, the heat energy entering the material increases the vibration of the particles/atoms at that end. These atoms also set the other neighboring atoms into vibration transferring the energy to the other end.

b) Free/mobile/delocalized electrons

Generally metals have free electrons which move all over the metal body. Heat energy supplied at one end of the material increases the kinetic energy of the electrons which then transfers this energy to the other end.

The ability of a material to conduct heat is called thermal conductivity. Materials that easily conduct heat are referred to as good conductors e.g. most metals. This is because metals conduct both by atom vibration as well as movement of free/delocalized electrons. This is why cooking utensils, soldering irons are made from metals. Non-metals are generally bad conductors of heat because they do not have free electrons.

The set up below can be used to compare thermal conductivity of different materials:



The order in which wax on the rods fall is the order of thermal conductivity of the materials i.e. the rod from which wax first falls is the best thermal conductor and decreases in that order. It is important to ensure the following when doing this comparison:

- Use same length of the rods with the same length in the hot water.
- Keep the temperature of the water bath uniform
- Use same thickness or amounts of wax

Below is a table of good and bad conductors:

Good thermal conductors	Bad thermal conductors
Silver	Concrete
Copper	Glass
Aluminum	Brick
Brass	Asbestos paper
Zinc	Rubber
Iron	Wood
Lead	Water
Mercury	Air

NB:

- During thermal conduction, heat flows through materials without the materials shifting or moving.
- Thermal conduction requires a material medium.

Factors affecting thermal conduction in solids

- **Temperature difference between the ends of the material**

More heat flows when the temperature difference is large.

- **Cross-section area**

The thermal conductivity of a thicker material is higher than that of a thinner one of the same material. This is because the number of free electrons (metals only) per unit length of the thicker material is higher than that of a thinner one. Also, the number of atoms vibrating per unit length of the thicker material is higher than that of a thinner one.

- **Length of the conductor**

Heat energy reaches the other end of a shorter conductor faster than a longer one. Heat travels in a conductor along imaginary lines known as lines of heat flow. These lines diverge away from the hot end and therefore as the length of the conductor increases, some of the energy is 'dropped'.

- **Nature of the material**

Some materials are good conductors while others are bad conductors e.g. copper is a better conductor compared to iron.

The above four factors can be combined together in the following equation:

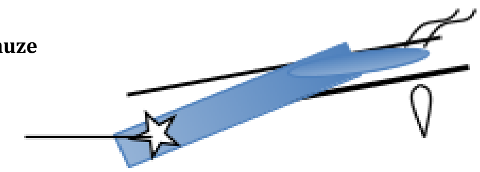
$$\textbf{Rate of heat flow} = \frac{\textbf{thermal conductivity of the material, } k * \textbf{Cross-section area, } A * \textbf{Temp. Difference, } \Delta\theta}{\textbf{Length, } l}$$

Lagging

This involves covering good conductors with bad one (insulators) with the aim of minimizing heat loss to the surroundings. For instance pipes used to convey hot water from boilers are normally covered using thick asbestos material.

Conduction in liquids

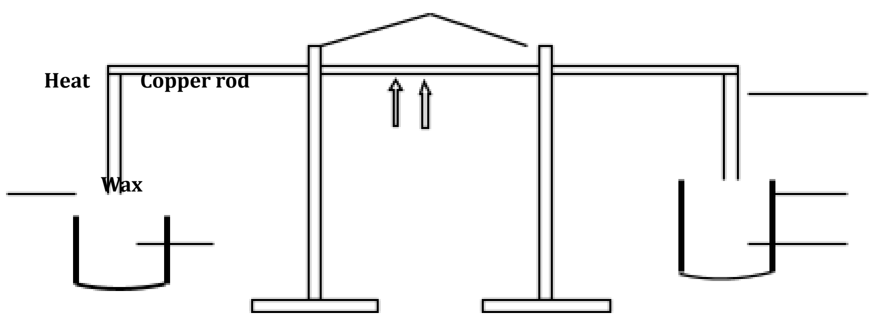
Conduction in **liquids** is not as pronounced as in solids. Liquids are generally poor thermal conductors with the exception of a few like mercury and some electrolytes (e.g. salt solution). Note that mercury exists as a liquid at room temperature. Thermal conductivity of water may be investigated by the set up below:



Ice wrapped in wire gauze is placed at the bottom of a boiling tube containing water and heated from the top. It is observed that as water at the boils producing steam, the ice remains unmelted. This shows that heat did not reach the bottom even having been aided by the wire gauze which is a good conductor. In this experiment:

- The boiling tube is made of glass which is a poor thermal conductor and thus limits the possible heat conduction down the tube by glass.
- The ice is wrapped in wire gauze to ensure that it does not float on water. Wire gauze is a good thermal conductor but still the ice did not melt, indicating that there is very little, if any, heat conducted down by water.
- Heating water at the top eliminates the possibility of heat transfer to the ice by convection.

Although liquids are poor thermal conductors, some liquids are better conductors than others. This can be investigated by the set up below:

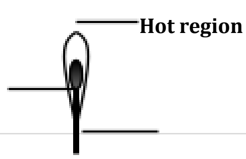


The test tubes are coated with a uniform layer of wax. When the rod is heated as shown, the wax on the test tube with mercury is observed to melt downwards after sometime. Later on, wax on the other test tube started melting but very slowly.

In conclusion, mercury is a better thermal conductor compared to water. Note that the heating should be done at the centre for a fair result. Also, the region of the rods in the liquids should be the same.

Conduction in gases

Gases are the worst thermal conductors. This is because they have larger intermolecular distance minimizing collision between their molecules. This can be verified by the set up below:



Match stick

A match stick held in the unburnt gas region of a flame is never ignited by the heat from the hot region above it. This shows that gases are poor thermal conductors.

Applications of good and poor conductors

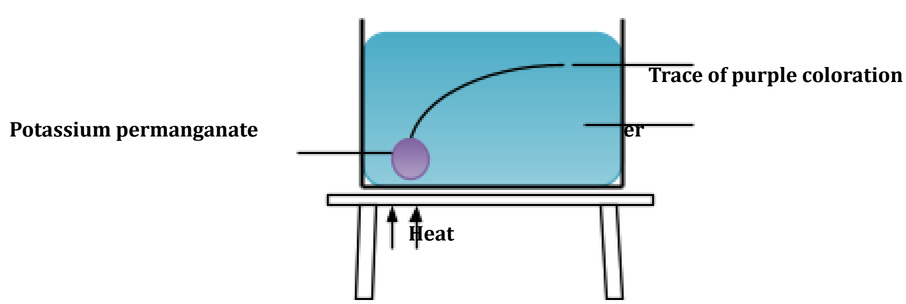
- Cooking utensils, soldering iron and boilers are made of metals because they are good thermal conductors. In some cooking utensils, their handles are covered using bad heat conductors such as wood, plastic
- Integrated circuits (ICs) and transistors in electronic devices are normally fixed to a heat sink; a metal plate with fins. This is to conduct away undesired heat which may otherwise affect the operation of the device.
- Fire fighters put on protective clothing made from asbestos material to keep them safe while putting out fires. Film actors involved in stunts involving burning also put on similar clothing.
- Birds flap their wings after getting wet in order to introduce air pockets in their feathers. Air is a poor thermal conductor and hence reduces heat loss from their bodies. The same applies to wool, fur, and thatch on roofs. Soft board ceiling have many air pockets than concrete one hence its preference to the concrete ceiling.
- When heating liquids using a glass beaker, the beaker is usually placed on a wire gauze. A wire gauze is a good heat conductor and hence spreads out the heat to a larger area of the beaker. If the gauze was not used, heat would only be concentrated on a small area and the beaker may crack as a result.
- In some buildings where the inside temperature is to be stabilized, double walls are constructed with an insulator like glass wool or foam plastic between them. Also, double glazed windows have air trapped between the two glass sheets.

7.2.2: Convection

It is a process by which heat is transferred in fluids. The term fluid refers to both liquids and gases. In this case heat transfer is by actual movement of the fluid itself. When a portion of liquid at the bottom of a container is heated, the particles there acquire more energy and expand. Since its mass does not change, the liquid in that portion becomes less dense and rises to the top. Cold, heavy portion of the liquid move down to replace the warm rising portion. A circulation of hot and cold water sets up a convectional current. Heat transfer in this manner is called convection.

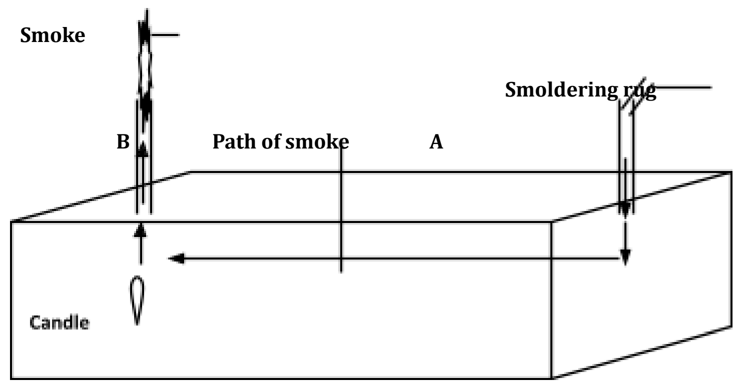
However, if heated from top no convection current is set up. This is because hot, lighter liquid cannot sink down below the cold, heavy liquid.

To show convection in **liquids**, a beaker is filled with water and some crystals of potassium permanganate placed at one corner of the beaker.



When heated from the position shown, a purple coloration is observed to rise from the potassium permanganate crystal forming a loop. Hence when a liquid is heated, it rises and a cold one descends to replace it. This movement of the liquid forms convection currents.

Convection in gases can be demonstrated using the set up below:

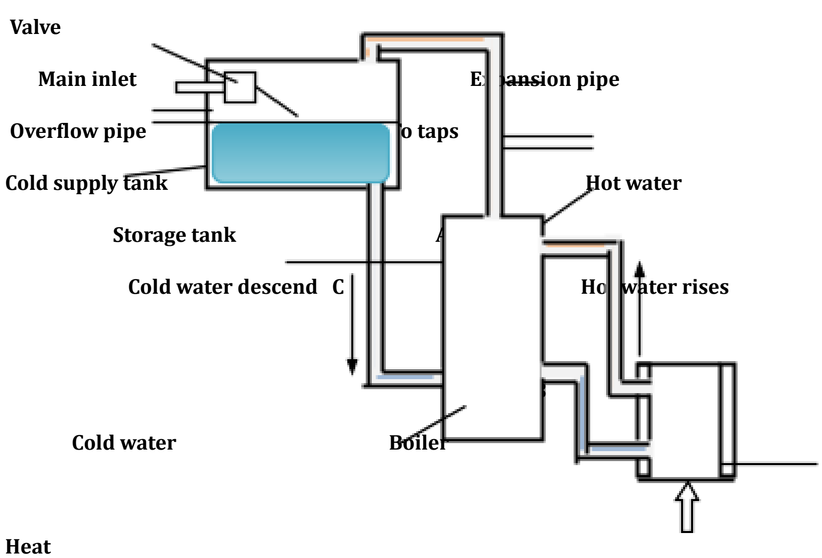


A smoldering straw/ rug burning on one end is used to introduce smoke through chimney A into the box. After sometime, the smoke is observed to exist through chimney B. however, when the candle is put off, NO smoke is sucked into the box.

The candle heats up the air above it which then expands and rises up chimney B. cold , heavy air is drawn in through chimney A to replace the rising warm air. It carries along with it smoke particles. Hence smoke exists through chimney A so long as the candle is burning and the produces smoke.

Applications of convection

I. **The domestic hot water system**



Water is heated up in the boiler, expands thereby becoming less dense. It then rises up tube A. cold water descends to the boiler through the return pipe B. The force of gravity helps in pushing the cold water down from the cold water tank to the boiler. The storage tank thus contains hot water on the upper part and cold water on the lower part. The valve controls the amount of water in the cold water tank. The overflow pipe helps in removing excess water from the cold water tank.

The pipes conveying hot water, the boiler and the storage tank are normally lagged to minimize heat loss through conduction.

II. House ventilation

Generally windows of houses are normally put closer to the floor and ventilation holes/openings high up in the walls. Air expelled by the house occupants is warm, lighter and thus rises and escapes through the ventilation openings. Cold, fresh heavier air flows into the room through the windows and doors to replace the escaping air. Thus a continuous circulation of air is set up in the room.

In certain modern houses, air conditioning devices are fitted to bring about forced convection of air, giving out cold dry air and absorbing warm moist air.

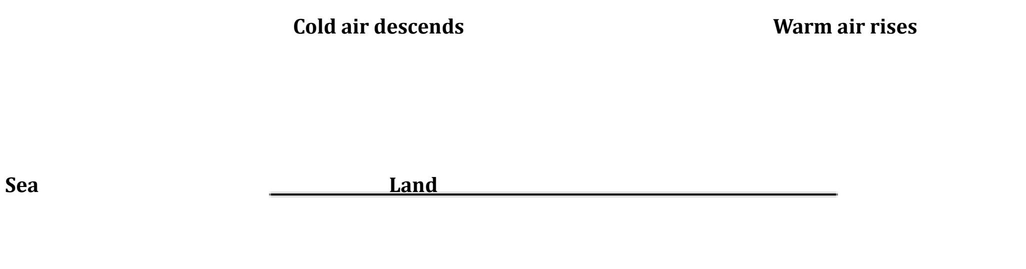
III. Car engine cooling system

A car engine gets heated as it operates. This heat must be disposed of to ensure the engine remains efficient. The engine is normally surrounded by a metallic water jacket connected to the radiator. The metal surface conducts away heat from the engine to the water. Water is heated up causing convection currents. The hot water is pumped into the radiator with thin copper fins that further conducts the heat away from the water. Fast moving air flowing between the fins speeds up the cooling process.

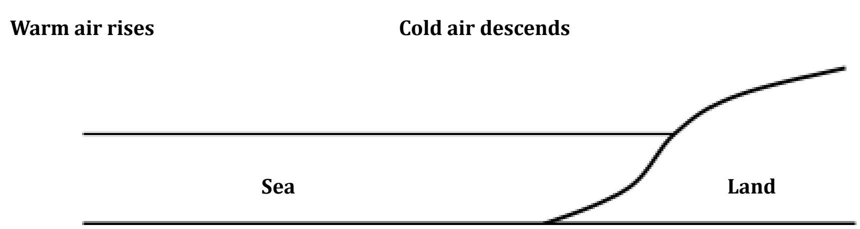
IV. Land and sea breeze.

This is a natural occurrence. A mass of water takes a much longer time to be heated up to the same temperature as a nearby land. It also loses heat slowly compared to land. It is on this argument that land and sea breezes are formed.

During the day, land gets heated much faster than a nearby mass of water. The mass of air just above the land is thus heated faster, expands and rises due to its reduced density. Cold air from above the water e.g. sea, drifts towards the land to replace the warm rising air. This is termed as sea breeze.

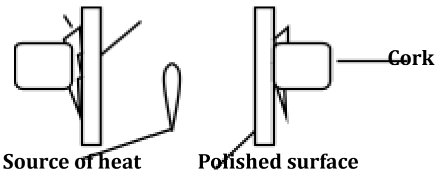


At night, the land cools faster than the sea. The air above the sea will be at a higher temperature. It expands and rises up due to its low density. Cold air from the land drifts towards the sea to replace the warm rising air. This constitutes a **land breeze**.



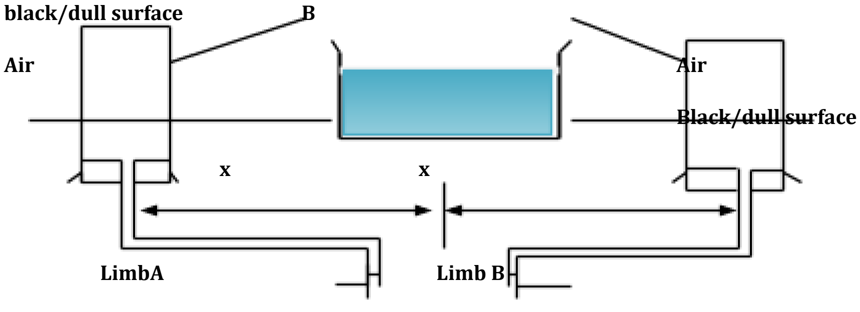
7.2.3: Radiation

It is the flow of heat energy from one point to another by means of electromagnetic waves. It requires NO material medium. These waves are able to pass through a vacuum. Energy from the sun reaches the earth's surface by radiation. All bodies emit and absorb radiant energy (radiation). A body emitting thermal radiation can also emit visible light when it is hot enough. The higher the temperature of a body the greater the amount of radiant energy it can emit. Thermal radiation can be reflected and causes a heating effect. The amount of radiant energy absorbed by a body depends on the nature of its surface. Dull/ black surfaces are good absorbers than shinny/ polished surfaces. This can be illustrated by the set up below:



After some time, the cork on the dull surface is observed to fall off first once the wax melts. The one on the polished surface takes longer to fall off. This shows that dull surfaces are good absorbers of radiation than polished surfaces. The radiation falling on the polished surface is partly absorbed and partly reflected.

Generally, good heat absorbers are also good emitters/ radiators. Hence black/ dull surfaces are good emitters while polished surfaces are poor emitters. Note that poor emitters (polished surfaces) are good reflectors.



After some time, the level of water in limb A is observed to rise while that in limb B falls. The boiling tube B receives more heat than A. Hence the air in B expands pushing the water in limb B down. This is a proof that dull surfaces are good emitters.

Applications of thermal radiators

1. Kettles, cooking pans and iron boxes have polished surfaces to minimize heat loss through radiation.

2. Petrol tanks are painted silvery bright to reflect away as much heat as possible.

3. Buildings which are whitewashed or painted using bright colours remain cool even during hot conditions because the bright colours reflect much of the sun's radiation. In cold regions, it is the inner walls and roofs which are painted with bright colours and the outer walls with dull colours. In very hot countries, people do put on white clothing because they reflect much light.

4. **Solar concentrators**

Concave reflectors concentrate light that fall on them to a common point. The temperature at this point may be high enough to boil water.

5. Green house effect

A green house acts as a heat trap by preventing the reflected radiation from the earth's surface from passing through glass. High energetic short wavelength radiation from the sun passes through glass without being absorbed. Inside the greenhouse, this heat is absorbed by the earth and the objects in it. These objects in turn emit long wavelength radiation of low amount of energy that cannot penetrate glass. This raises the temperature inside the greenhouse. This is called greenhouse effect.

Greenhouses provide optimum conditions for plants in cold regions. In the atmosphere especially at the lower layers, the gases and air pollutants suspended behave like glass, trapping the long wavelength reflected radiation. This may raise the temperature of the atmosphere to some dangerous levels, a phenomenon known as global warming.

6. Solar heater

A solar heater uses radiation from the sun to heat water. It comprises a coiled blackened copper pipe resting on a blackened insulating surface. Radiation from the sun passes through glass and absorbed by the copper pipes. Copper conducts this heat to the water flowing through it.

7. A thermos (vacuum) flask.

A vacuum flask is designed to keep its contents at a constant temperature by minimizing heat transfer into or out of the flask. It has a double wall; silvered to prevent heat loss or heat gain through radiation. Between the two walls is a vacuum which is a perfect insulator and minimizes heat loss through conduction and convection. The cork/lid minimizes heat loss through convection.

TOPIC 8: RECTILINEAR PROPAGATION OF LIGHT AND REFLECTION AT PLANE SURFACES

8.1: Light

Light is a form of energy that enhances vision. Certain objects are able to produce their own light and are called luminous or incandescent objects. They include the sun, stars, burning candle, etc. however, most objects cannot produce their own light and can only be seen when light falling on them is reflected. Such objects are referred to as non-luminous objects. They include the moon, planets, plants, walls, clothes, etc.

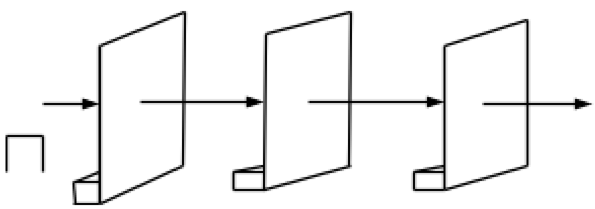
The path along which light energy travels is called **a ray** of light. It is usually represented by a line with an arrow to show the direction of travel. A group or bundle of light rays is called a **beam** of light.

Different objects behave differently when light fall on them. For instance, some objects do not allow light to pass through them totally. Such objects are said to be **opaque**, e.g. brick walls, metals, wood, etc. some allow light through them and one can see through them. These are said to be **transparent** and they include ordinary car windscreen, normal glass window panes. Some on the other hand allow light to pass through them but one cannot see through them. They are called **translucent** objects and they include greased paper, some glass panes, etc.

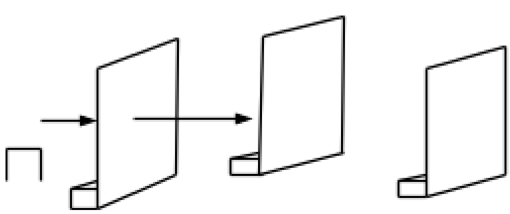
8.2: Rectilinear propagation of light.

This is a property of light which suggests that light travels in a straight line. To investigate this, three cardboards each with a hole at its centre are arranged such that the holes are along a straight line. Light from a lit candle on one side of the cardboards is seen from the other end as shown in (a) below. However, when one of the cardboards is displaced, light is cut off as shown in (b).

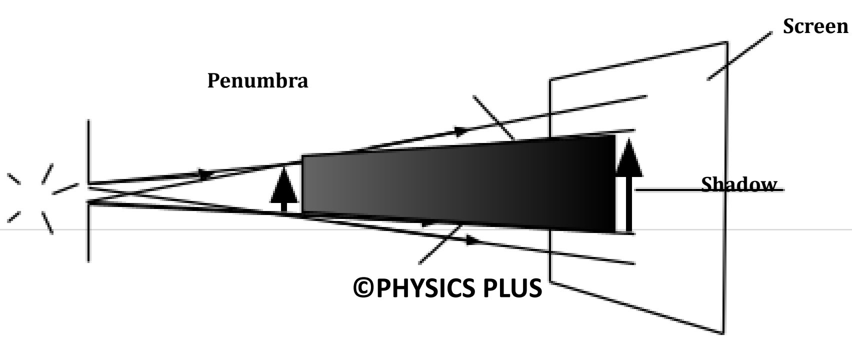
a)



b)



Another way of demonstrating rectilinear propagation of light is by placing an opaque object along the path of light. A shadow is formed whose nature depends on the size of light source, size of the object and the distance between the distance between the object and the source of light.



Object

Umbra

When the light source is small e.g. a point source, the shadow formed is uniformly dark and is called umbra. However, when the light source is large, the shadow formed will have a partial shadow along its edges. This is called penumbra. Note that points within the umbra receive **NO** light at all while points within the penumbra receive some amount of light but not as much as it would if the obstacle was not there.

8.3: Eclipses

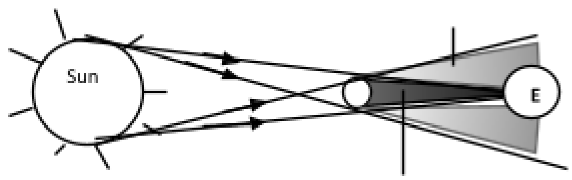
An eclipse occurs when there is partial or total disappearance of either the sun or moon when viewed from the earth. There are three types of eclipses namely; the solar eclipse, annular eclipse and lunar eclipse.

8.3.1: Solar eclipse

It is also called eclipse of the sun. It occurs when the moon passes between the sun and the earth. Since the sun is bigger than the moon, both the umbra and penumbra are formed.

Penumbra

Earth



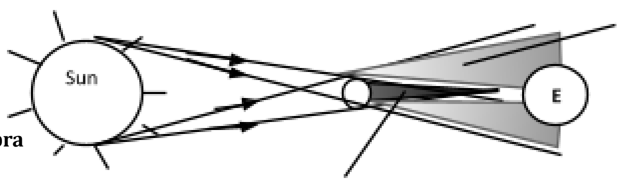
Umbra

Sometimes the tip of the moon umbra fails to reach the surface of the earth. This is called an annular eclipse.

Moon **Penumbra**

Earth

Umbra



Note that solar eclipse is very rare. This is because the Moon's orbit around Earth is inclined at an angle of just over 5^0 to the plane of Earth's orbit around the Sun (the ecliptic) i.e. the path of the moon around the earth is not a perfect circle.

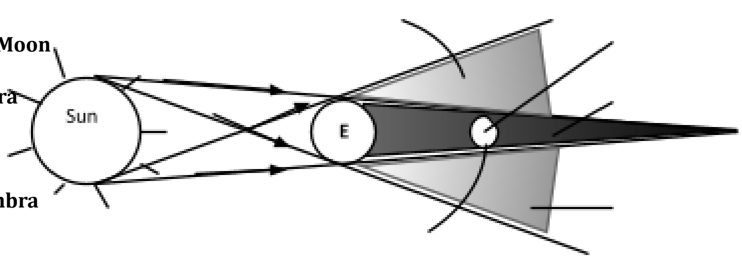
8.3.2: A lunar eclipse/eclipse of the moon

It occurs when the earth is between the moon and the sun. The shadow cast by the earth prevents light from the sun from reaching the moon i.e. during the lunar eclipse the moon passes through the earth's umbra. This eclipse occurs only at night when there is a full moon.

Earth Moon

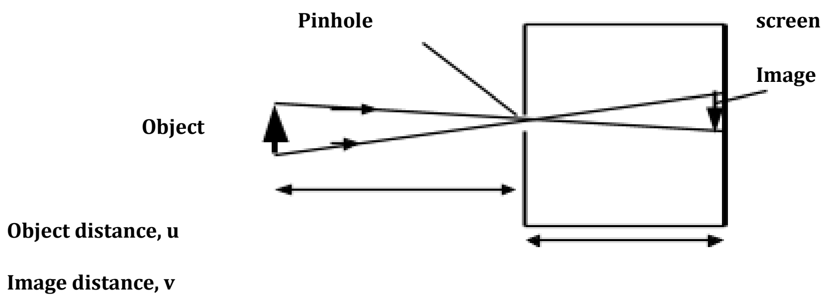
Umbra

Penumbra



8.4: The pin-hole camera

It is a closed box which is painted black on the inside. It has a small hole on one face and a screen made of a translucent paper on the opposite face. The object to be viewed is placed in front of the face with the hole.



The rays from the object fall on the screen forming an inverted image. The inner part of the camera is painted black so as to prevent reflection of the light. The size of the image formed depends on the distance of the object from the pinhole, u and the distance of the screen from the pinhole, v. if the camera is moved nearer the object, the image becomes bigger.

When the hole is made larger the image becomes blurred (not clear). This is because many rays of light will be allowed to the screen, each forming its own image. The result is overlapping images or blurred image.

Generally, a pinhole camera forms a real and inverted image.

One advantage of the pinhole camera is that it forms focused images of both far and near objects.

A pinhole camera has some **limitations** which include:

- It cannot take motion pictures.
- Its exposure time is too long due to the size of the hole.

The ratio of the image size to the object size is referred to as **magnification.**

i.e. magnification, m= image size/object size

It can also be shown that magnification is the ratio of the image distance to the object distance.

i.e. magnification, m= image distance, v/object distance, u

Magnification has no units.

Example 8.1

1. The distance between the pinhole and the screen in a pinhole camera is 20cm. a student uses the camera to form an image of a person 4m away. The person's height is 140cm. what is the height of the image? Determine the magnification.

$v/u = h_i/h_o$

h_i = (140cm*20cm)/400cm = 7.0cm

$m = v/u = 20/400 = 0.05$

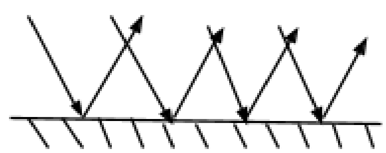
2. A lamp of height 6cm stands in front of a pinhole camera at a distance of 24cm. the camera screen is 8cm from the pinhole. What is the height of the image of the lamp on the screen?

   $v/u = h_i/h_o$

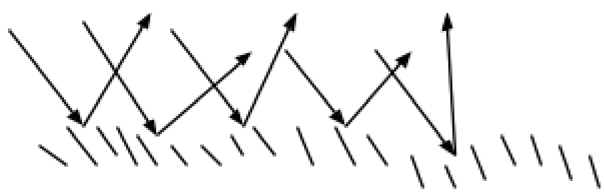
$h_i = (6cm*8cm)/24cm = 2cm.$

8.5: Reflection of light

When a ray of light strikes a surface, part of it bounces off. When the surface is smooth or highly polished, all incident light is reflected back uniformly. This is called **regular/specular reflection.**



A rough surface on the hand results in an **irregular/diffuse reflection.**



8.5.1: Terms used

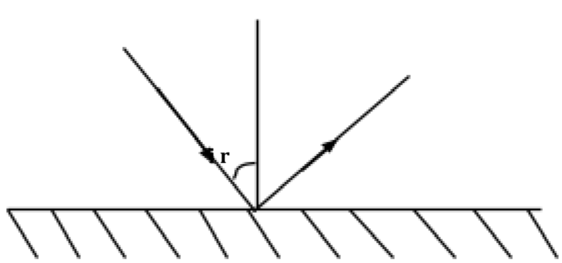
Incident ray- is a ray travelling from the source to the reflecting surface.

Reflected ray- is the ray that bounces off from the reflecting surface.

Normal- is a line drawn perpendicularly at the point where the incident ray strikes the reflecting surface.

Angle of incidence- is the angle between the incident ray and the normal.

Angle of reflection- is the angle between the reflected ray and the normal.



8.5.2: LAWS OF REFLECTION

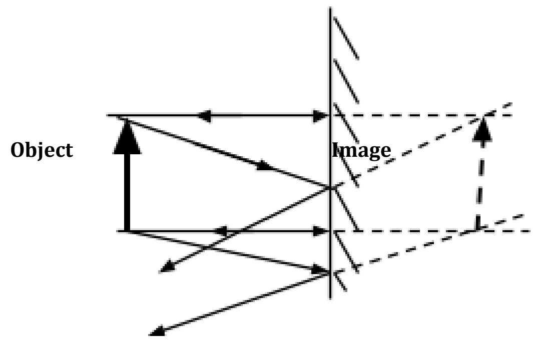
There are two laws of reflection of light:

i. The angle of incidence and the angle of reflection are equal.

ii. The incident ray, reflected ray and the normal at the point of incidence all lie on the same plane.

8.6: Images formed by a plane mirror.

The figure below shows how a plane mirror forms images of objects placed in front of them:



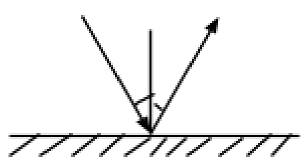
Generally, an image formed by a plane mirror has the following characteristics:

- It is virtual i.e. not formed by actual intersection of real rays.
- It is erect/upright.
- It is the same size as the object.
- It is laterally inverted i.e. the sideway turning effect.
- It is the same distance behind the mirror as the object is in front of the mirror i.e. image distance = object distance.

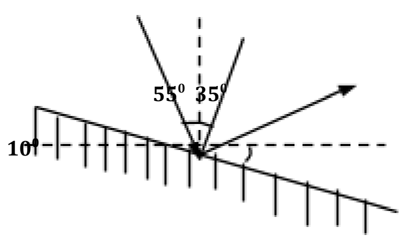
8.6.1: Rotation of a mirror through an angle

Consider a ray incident on a plane mirror as shown below. The angle of incidence is 25^0.

25^0

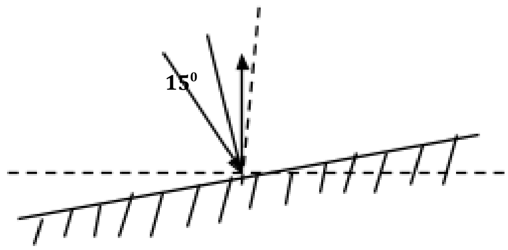


Rotating the mirror through 10^0 about the point of incidence in the clockwise direction, the result is as shown in below:



The new angle reflection is 35^0 from 25^0.

When the mirror is rotated through 10^0 about the point of incidence in the anticlockwise direction, the result is as shown below:



The new angle of reflection is 15^0.

8.6.2: Mirrors at an angle

The number of images formed by two plane mirrors placed at an angle to each other depends on the angle of inclination. The number of images formed by such an arrangement can be calculated from the formula; **$n= 360^0/\theta - 1$**

Where n- the number of images formed

Θ- Angle of inclination of the mirrors.

Example 8.2

1. At what angle should two mirrors be inclined to form:

a) 17 images?

$17= 360/\theta -1$

$18*\theta=360$

$\Theta=360/18 =20^0$

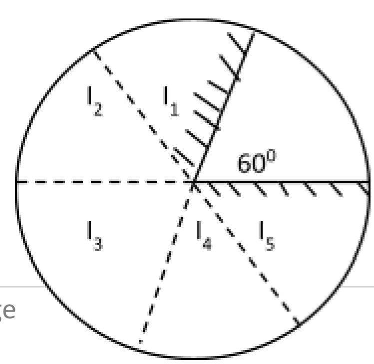
b) 29 images?

$\Theta=360/30 =12^0$

8.7: Applications of reflection at plane surfaces

1. The kaleidoscope

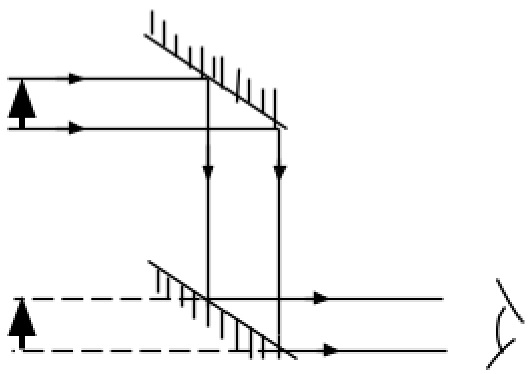
It comprises of two plane mirrors inclined at an angle of 60^0 inside a tube. It has a ground-glass plate to admit light. A piece of brightly coloured glass is placed on the glass plate. This acts as an object. When one looks down the tube, five images of the object are seen which together with the object form a symmetrical pattern in six sectors.



A kaleidoscope may be used by designers to obtain ideas on systematic patterns.

2. The periscope

This is an object that can be used to view objects over obstacles. It consists of two plane mirrors facing each other at an angle of 45^0. It is used to in submarines and also to watch over crowds.



A periscope in general forms erect (upright) and virtual images. Periscopes used in submarines are more elaborate, in which prisms are used instead of plane mirrors and the tube supporting them incorporates a telescope to extend the range of vision.

TOPIC 9: ELECTROSTATICS I

9.1: introduction

Electrostatics refers to the study of static charges. Some of the cases illustrating the effects of static charges include:

a) When a plastic ruler or pen is rubbed on the hair, it picks up small pieces of paper when it is brought closer to them.
b) When a glass window pane is wiped using a dry piece of cloth on a dry day, it immediately attracts dust particles.
c) When a nylon cloth is taken off the body, a cracking sound is produced.

The above observations result from the formation of static charges. These charges are as a result of friction between the rubbed surfaces. Generally when a glass rod is rubbed using silk, it gains positive charges while when a polythene rod is rubbed using fur or cloth, it gains negative charges.

There are two types of charges namely negative and positive charge. When the number of negative and positive charges in any material are the same then the material is said to be neutral. When there are more positive charges than negative charges, the body is said to be positively charged while if there are more negative charges than the positive charges, the body is said to be negatively charged.

The origin of charge can be traced back to the atom. An atom consists of smaller particles namely protons (positively charged), electrons (negatively charged) and neutrons (neutral in nature- has no charge). Protons and neutrons are found at the nucleus while electrons are found on the energy levels around the nucleus. For any atom, the number of protons equal to the number of electrons. Hence an atom is neutral.

9.2: Basic law of electrostatics

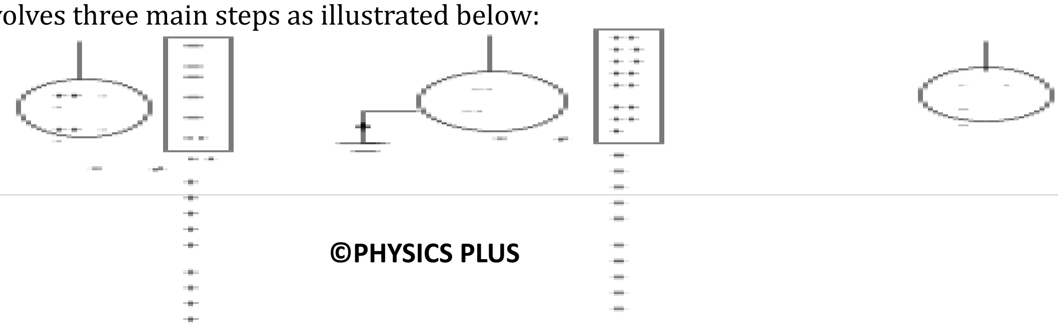
This law states that unlike charges attract while like charges repel.

9.3: charging a material

There are three ways of charging a body namely by induction, contact method or by separation.

9.3.1: Induction method

This method involves three main steps as illustrated below:



e^-

i) ii) iii)

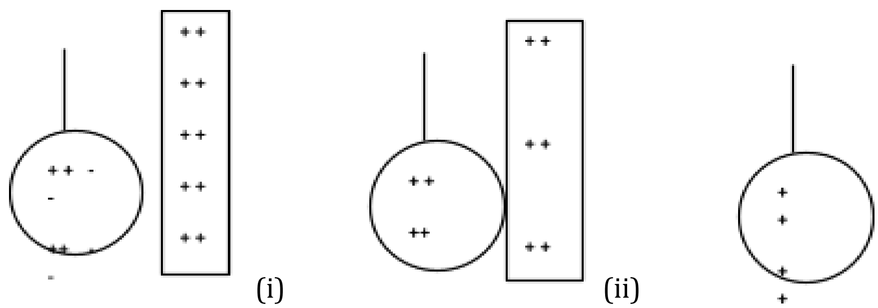
In step (i), a positively charged glass rod is brought close to a suspended polystyrene ball. The negative charges will be attracted towards the side with the glass rod while the positive charges are repelled away.

The ball is then earthed by touching it with the finger on the side away from the glass rod as shown in (ii) above. Electrons will flow from the earth to neutralize the positive charges on the ball.

In step (iii), the finger is withdrawn first and then the glass rod. The negative charges repel each other thereby spreading all over the ball. Hence the polystyrene ball becomes negatively charged.

Note that when a body is charged by induction method, it acquires an opposite charge to that of the inducing charge.

9.3.2: Contact method



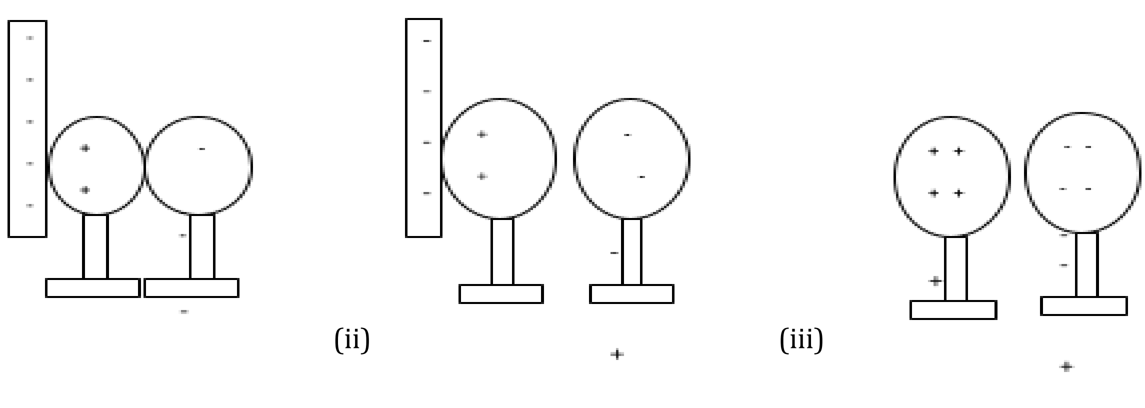
(i) (ii) (iii)

A positively charged glass rod is brought into contact with a suspended polystyrene ball and then withdrawn. When the glass rod is rolled on the polystyrene ball, the induced negative charges on the ball are neutralized. When the glass rod is withdrawn, the positive charges are redistributed all over the ball. Hence the ball becomes positively charged.

Note that when a body is charged by contact method, it gains the same charge as the charging material.

9.3.3: Separation method

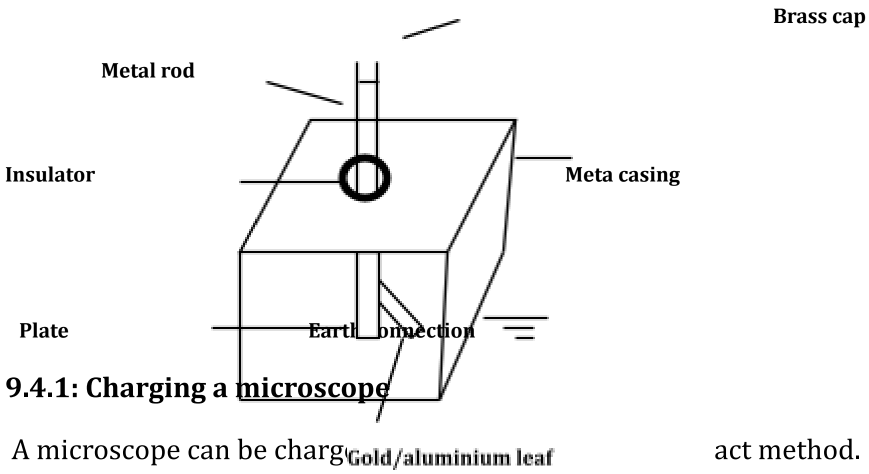
Two spheres are placed in contact so that they form a single conductor. A negatively charged rod is then brought close to but not touching one of the spheres as shown below in figure (i) below.



(i) (ii) (iii)

With the rod in position, the contact between the two spheres is broken as in (ii) above. When the rod is finally withdrawn, the charges on the individual spheres are redistributed and they eventually acquire opposite charges as shown in (iii) above.

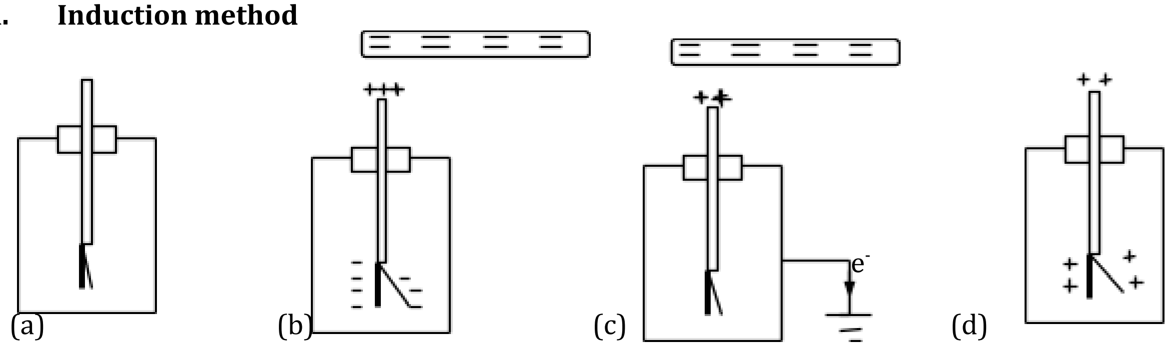
9.4: The gold-leaf electroscope



9.4.1: Charging a microscope

A microscope can be charg... act method.

i. **Induction method**



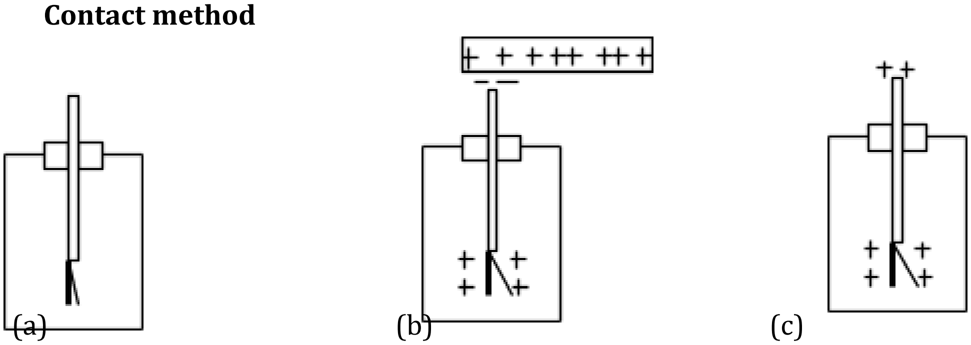
When a negatively charged polythene rod is brought near the cap of uncharged electroscope, positive charges are attracted to the cap leaving negative charges on the plate and leaf. The leaf thus diverges.

With the rod in position, the electroscope is earthed by touching the cap. Electrons then flow from the electroscope to the earth. The leaf momentarily falls but when the finger and the rod are withdrawn in that order, the positive charges are redistributed throughout the electroscope and the leaf diverges again.

Assignment 9.1

1. Describe how to charge an electroscope negatively by induction method.

ii. **Contact method**



The negative charges on the electroscope are attracted to the cap and neutralized leaving positive charges on the plate and leaf. Hence the leaf diverges. When the rod is withdrawn the positive charges are redistributed all over the electroscope.

Note that when an electroscope is charged by contact method, it acquires the same charge as that of the charging material.

Assignment 9.2

1. Describe how to charge an electroscope negatively by contact method.

9.4.3: Uses of the electroscope

a) To distinguish between a conductor and an insulator.

When a conductor is brought into contact with a charged electroscope, the leaf falls. When the electroscope is positively charged, electrons flow from the earth through the material to the electroscope to neutralize the positive charges. When the electroscope is negatively charged, electrons flow through the material to the earth. Hence the leaf falls.

However, for an insulator there will be no effect on the leaf divergence of the electroscope.

b) To test the quantity of charge on a charged body.

The degree of leaf divergence is proportional to the quantity of charge on a body i.e. the higher the quantity the larger the divergence.

c) To detect the presence of charge on a body.

When a charged body is brought close to the cap of an electroscope, the behavior of the leaf will determine whether the body is charged or not. This will be looked at in the next point. However, when an uncharged body is brought close to the cap of a charged electroscope; either positively or negatively, the leaf divergence reduces while if the electroscope is uncharged, they will be no effect on the leaf divergence.

d) To test the sign of the charge on a charged body.

In this case, the initial charge on the electroscope must be known. When a body having an opposite charge is brought near the cap of the electroscope, the leaf divergence decreases due to attraction.When a similar charge is brought near the cap, the leaf divergence increases due to repulsion. This is summarized in the table below:

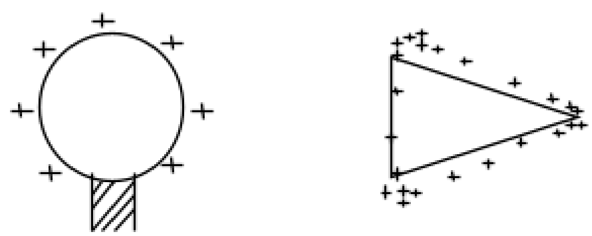
Charge on electroscope	Charge on the body	Effect on leaf divergence
+	+	Increase
-	-	Increase
+	-	Decrease
-	+	Decrease
+/-	Uncharged	Decrease

FORM THREE

ELECTROSTATICS II

9.5: Charge distribution on the surface of a conductor(FORM THREE WK)

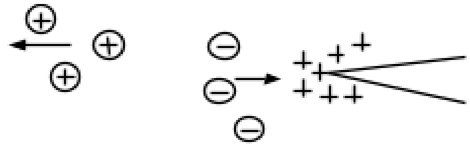
The quantity of charge per unit area of the surface of a conductor is called **charge density.** The charge distribution on a conductor depends on the shape of the conductor. Generally, the charge concentration on a spherical conductor is uniform while that on a sharp point is high.



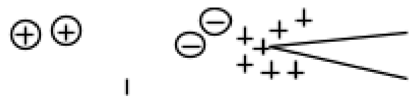
The high charge concentration at sharp points makes it easier to gain or lose charges. The effects of high charge concentration at sharp points can be seen in the following cases:

9.5.1: Electric wind

When a highly charged sharp point is brought close to a candle flame, the flame is observed to drift away as if there was wind. The high charge concentration at the sharp point ionizes the surrounding air producing both positive and negative charges. Opposite charges are attracted to the point while similar charges are repelled away from the point blowing away the flame.



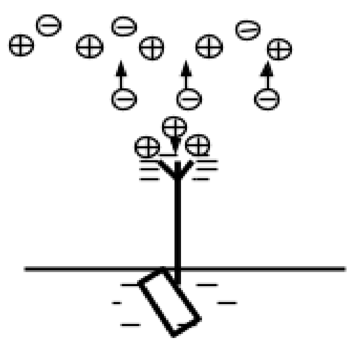
If the point is brought very close, the flame splits into two; one part moves towards the point and the other part away from the point. This is because a flame has both positive and negative ions. The negative ions are attracted towards the point while the positive ions are repelled away from the point.



9.5.2: Lightning arrestors

When clouds move in the atmosphere, they rub against the air particles and produce a large amount of static charges by friction. These charges induce large amounts of the opposite charge on the earth. Hence a high potential difference is created between the earth and cloud. This makes air to be a charge conductor. The opposite charges attract each other and neutralize, causing thunder and lightning. Lightning can be very destructive to buildings and other structures.

Lightning arrestors are used to safeguard such structures. It consists of a thick copper plate buried deep under the ground. The plate is connected by a thick copper wire to the spikes at the top of the building. The arrestor assumes the same charge as the earth. At the spikes, a high charge density builds up and a strong electric field develops between the cloud and the spikes. The air around the spikes is ionized. The opposite charges attract each other and neutralize. Excess electrons flow to the ground through the thick copper wire.



It is for this reason that people are advised not to take shelter under trees when it is raining.

9.6: Applications of static charges(JUST MENTION THE TWO)

- **Electrostatic precipitator**

One of the causes of air pollution globally is increased industrialization. Some industries have indeed responded to this challenge by installing electrostatic precipitators which are found within the chimneys.

An electrostatic precipitator consists of a cylindrical metal plate fixed along the walls of the chimney and a wire mesh suspended through the middle.

The plate is charged positively by connecting it to a high voltage, approximately 50,000V and the wire mesh charged negatively. As a result, a strong electric field exists between the plate and the wire mesh. The ionized pollutant particles get attracted; some to the plate and others to the wire mesh.The deposits are removed occasionally. The same principle is used in fingerprinting and photocopying.

- **Spray painting**

The nozzle of the spraying can is charged. When spraying, the paint droplets acquire similar charge and spread out finely due to repulsion. As the droplets approach a metallic body, they induce opposite charge which then attracts them to the metal surface. This ensures that little paint is used.

9.7: Dangers of static charges

When a liquid flows through a pipe, its molecules rub against each other and against the walls of the pipe and become charged. If the liquid is flammable like petrol, it is likely to cause sparks or even explosion. This can also happen to fuels when they are packed in plastic containers.

It is therefore advisable to store fuels and other flammable liquids in metallic containers so that any charges generated can continually leak out. This also explains why long chains hang underneath fuel tankers as they move.

9.8: Electric field(FORM THREE WK)

This is the region around a charged body where its influence (attraction and repulsion) can be felt. It is represented lines of force called electric field lines. The direction of an electric field is the direction in which a positive charge would move if placed at that point.

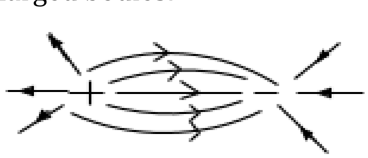
Electric field lines have the following properties:

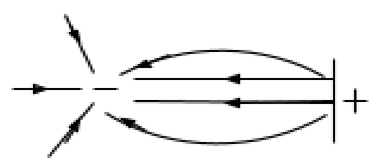
- Originate from a positive charge and terminate at a negative charge
- Do not cross each other i.e. do not intersect
- Are parallel at uniform field, close together at strong fields and widely spaced at weaker fields.

9.8.1: Electric field patterns

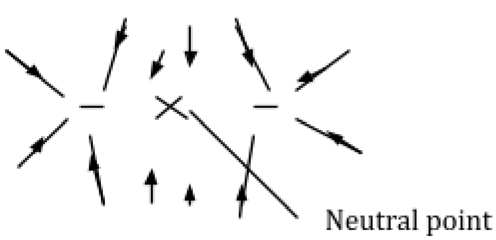
The electric field pattern between two charged bodies obeys the law of electrostatics. Below are some patterns between charged bodies:

(a) (b)





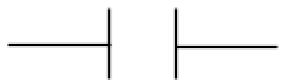
c)



NB/At the neutral point, the resultant effect is zero.

9.9: Capacitors

A capacitor is a device used for storing charge. It consists of two or more metal plates separated by a vacuum or a material medium (insulator). This material is known as a 'dielectric'. Other materials that can be used as a dielectric include air, plastic, glass e.t.c. the symbol of a capacitor is shown below:

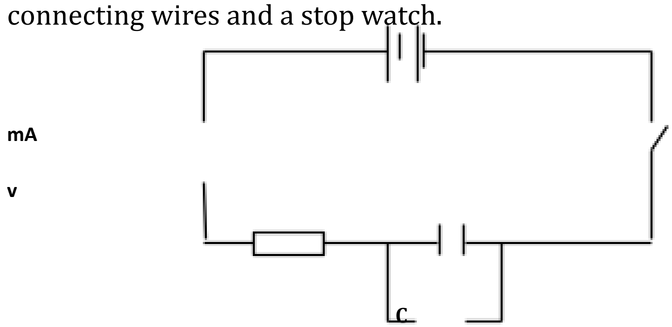


There are three main types of capacitors namely paper capacitors, electrolytic capacitors and variable capacitors. Others include plastic, ceramic and mica capacitors.

9.9.1: Charging a capacitor

Experiment: To charge a capacitor

Apparatus : Uncharged capacitor of 500μF, 6.0V power supply, rheostat, voltmeter, milliammeter, switch, connecting wires and a stop watch.



Procedure

i. Set up the apparatus as shown above.

ii. Close the switch and record the values of current, I at various time intervals. Tabulate your values in the table below:

Time, t(s)	**0**	**10**	**20**	**30**	**40**	**50**	**60**	**70**
Current, I(mA)								
It (mAs)								

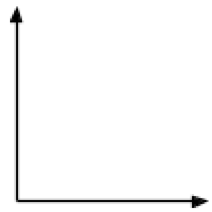
iii. Plot a graph of current, I against time, t

iv. Plot a graph of It against time.

Observations

The charging current is initially high but gradually reduces to zero. A graph of current, I against time appears as shown below:

I (mA)

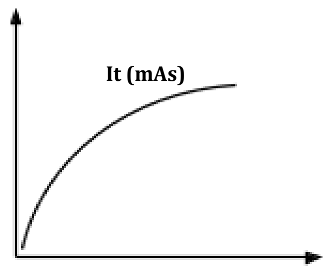


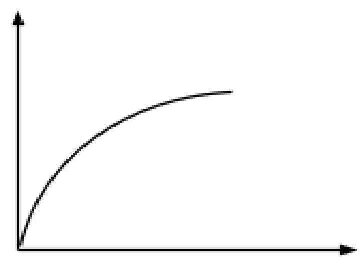
t (s)

The charging current drops to zero when the capacitor is fully charged. As the p.d. across the capacitor increases the charge in the capacitor also increases up to a certain value. When the capacitor is fully charged, the p.d across the capacitor will be equals the p.d of the source.

A graph of p.d across the capacitor against time is exponential. A graph of It against time is also exponential.

p.d (V)





t (s) t (s)

NB

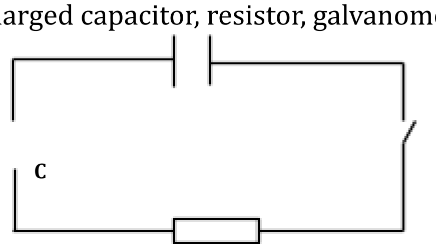
The product It represents the amount of charge in the capacitor.

9.9.2: Discharging a capacitor

Experiment: To discharge a capacitor

Apparatus :A charged capacitor, resistor, galvanometer, switch and connecting wires.

G



Procedure

- Set up the apparatus as shown above.
- Close the switch and record the values of current at various time intervals in the table below.

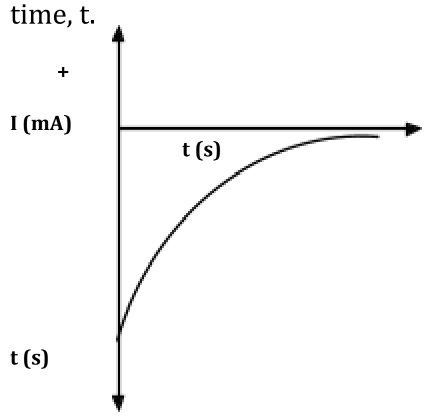
Time, t(s)	0	10	20	30	40	50	60	70
Current, I (mA)								

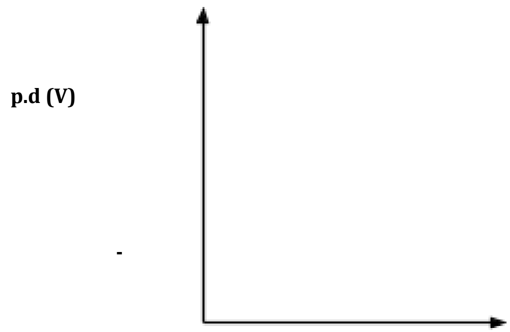
- Plot a graph of current, I against time, t.

Observations

The value of current is seen to reduce from maximum value to zero when the capacitor is fully discharged. The galvanometer deflects but in the opposite direction to that during charging.

During discharging, the p.d. across the capacitor reduces to zero when the capacitor is fully discharged. The graphs below show the variation between current, I and time, t and between the p.d across the capacitor and time, t.





A graph of charge in the capacitor, Q against time, t during discharging also appears like that of p.d against time i.e. p.d across the capacitor is directly proportional to the charge stored.

9.10: Capacitance

Capacitance of a capacitor is defined as the measure of the charge stored by the capacitor per unit voltage; $C = Q/V$

Hence $Q = CV$

Recall: $Q = It$

Therefore $Q= CV = It$

The SI Unit of capacitance is the farad, F. A farad is the capacitance of a body if a charge of one coulomb raises its potential by one volt.

Other smaller units of capacitance are: microfarad (μF), nanofarad (nF) and picofarad (Pf).

i.e. $1\ \mu F = 10^{-6}\ F$

$1\ nF = 10^{-9} F$

$1\ pF = 10^{-12} F$

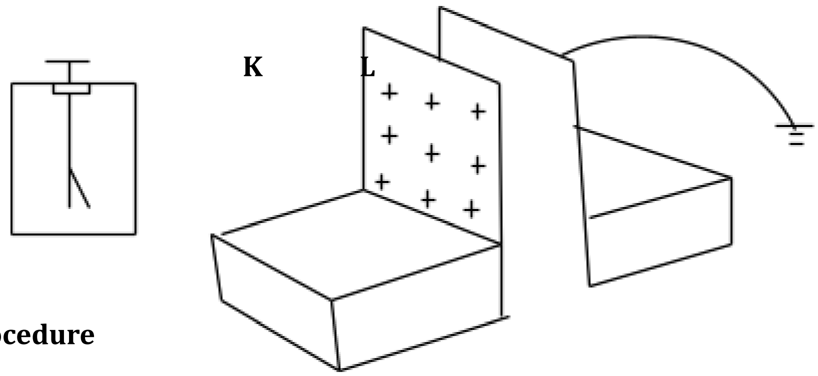
9.10.1: Factors affecting capacitance of a capacitor

The capacitance of a parallel plate capacitor depends on three factors, namely:

- Area of overlap of the plates, A
- Distance of separation, d between the plates
- Nature of the dielectric material

Experiment: To investigate the factors affecting capacitance

Apparatus: 2 aluminium plates, K and L of dimensions 25cm * 25cm**,**Insulating polythene support**,** uncharged electroscope**,** Glass plate**,** earthing wireand a free wire.



Procedure

- Fix the plates on the insulating support so that they stand parallel and close to each other as shown above.
- Charge plate K to a high voltage and then connect it to the uncharged electroscope. Earth the second plate, L.
- While keeping the area of overlap, A the same vary the distance of separation, d and observe the leaf divergence.
- While keeping the distance of separation, d constant vary the area of overlap, A and observe the leaf divergence.
- While keeping both the area of overlap and the distance of separation, d constant introduce the glass plate between the plates of the capacitor and observe what happens to the leaf.

Observations

1. When the distance of separation is increased the leaf divergence also increased.
2. When the area of overlap is increased the leaf divergence decreased.
3. When the glass plate is introduced between the plates, the leaf divergence increased.

Note that the leaf divergence here is a measure of the potential, V of plate K. Hence the larger the divergence the greater the potential and thus the lower the capacitance (since C = Q/V, but Q is constant).

Conclusion

From the above observations, it follows that the capacitance is directly proportional to the area of overlap between the plates and inversely proportional to the distance of separation. It also depends on the nature of the dielectric material.

$C \propto A/d$

$C = \varepsilon A/d$ where ε is a constant called permittivity of the dielectric material (epsilon).

If between the plates is a vacuum, then $\varepsilon = \varepsilon_0$, known as epsilon nought and is given by $8.85 * 10^{-12}$ Fm^{-1}. Hence **$C = \varepsilon_0 A/d$**

Example 9.1

1. How much charge is stored by a 300μF capacitor charged up to 12V? give your answer in (a) μC (b) C {**ans. 3600μC/0.0036C**}

   Solution

   a) Q= CV = 300 * 12 =3600μC b) 3600 * 10-6 =0.0036C

2. What is the average current that flows when a 720μF capacitor is charged to 10V in 0.03s?

   {ans. 0.24A}

   Solution

   Q = CV =It

   I= 720 * 10-6 *10 / 0.03 =0.24A.

3. Find the separation distance between two plates if the capacitance between them is 4.0 * 10^{-12}C and the enclosed area is 2.0 cm^2. Take ε_0 = 8.85 * $10^{-12}$$Fm^{-1}$. { d = 4.425 * 10^{-4} m}

   Solution

   C = ε_0A/d

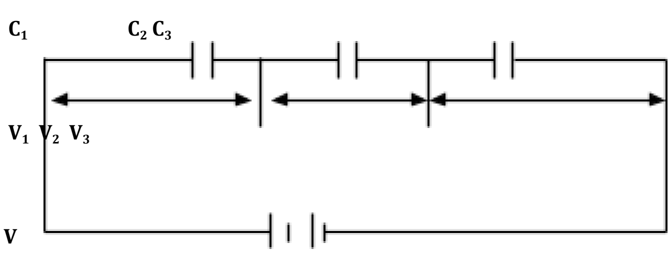
   d = 8.85 * 10^{-12} * 2.0 * 10^{-4} / 4.0 * 10^{-12}

   = 4.425 * 10^{-4} m

9.10.2: Arrangement of capacitors

a) Series arrangement

Consider three capacitors; C_1, C_2 and C_3 arranged as shown below:



Recall V = V1 + V2 + V3 and Q = CV

When capacitors are connected in series, the charged stored in them is the same and equals the charge in the circuit.

i.e. $Q = Q_1 = Q_2 = Q_3$

Therefore $V_1 = Q/C_1$, $V_2 = Q/C_2$, and $V_3 = Q/C_3$

$V = Q/C_1 + Q/C_2 + Q/C_3$

Dividing through by Q, we obtain $V/Q = 1/C_1 + 1/C_2 + 1/C_3$

Since $V/Q = 1/C$

$$\mathbf{1/C = 1/C_1 + 1/C_2 + 1/C_3}$$

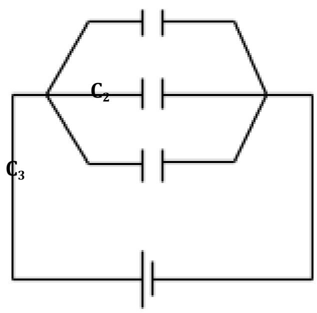
Where C is the combined capacitance.

In a special case of two capacitors in series, the effective/combined capacitance**,**

$$\mathbf{C = C_1C_2/ (C_1 + C_2).}$$

b) Capacitors in parallel

When capacitors are arranged in parallel, the potential drop across each of them is the same.



$Q_1 = C_1V$, $Q_2 = C_2V$, $Q_3 = C_3V$

The total charge, $Q = Q_1 + Q_2 + Q_3$

$Q = C_1V + C_2V + C_3V = V(C_1 + C_2 + C_3)$

Dividing through by V, we obtain $Q / V = C_1 + C_2 + C_3$

Since $C = Q/V$,

$$\mathbf{C = C_1 + C_2 + C_3}$$

Hence the combined capacitance for capacitors in parallel is the sum of their capacitance.

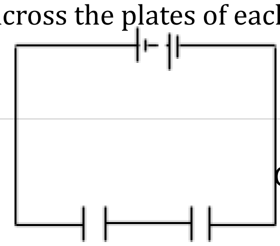
Example 9.2

1. In the circuit below, calculate:

a) The effective capacitance of the capacitors

b) The charge on each capacitor

c) The p.d across the plates of each capacitor



6V

12μF 24μF

Solution

a) $C = 12 * 24 / 12 + 24 = 8\mu F$

b) $Q_1 = Q_2 = CV = 8 * 6 = 48\mu C$

c) $V_1 = 48/12 = 4V$, $V_2 = 48/24 = 2V$

2. The figure below shows an arrangement of capacitors connected to a 10V d.c supply.

Determine: a) The combined capacitance of the arrangement

b) The total charge in the circuit

(ans. 0.7778μF,7.778μC)

a) $C_{BD} = 3*3/3+3 = 1.5\mu F$

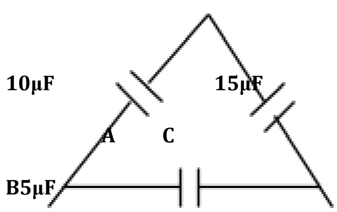
$C_{AE} = 2+1.5 = 3.5\mu F$

$C = 3.5*1/3.5+1 = 0.7778\mu F$

b) $Q = CV = 0.7778*10 = 7.778\mu C$.

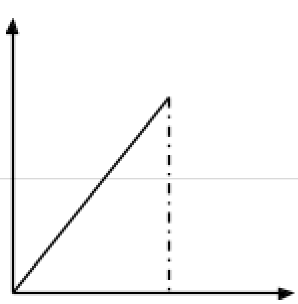
Assignment 9.3

The figure below shows part of a circuit connecting 3 capacitors. Determine the effective capacitance across AC.



9.10.3: Energy stored by a capacitor

During charging, the addition of electrons to the negatively charged plate involves doing work against the repulsive force. Also the removal of electrons from the positively charged plate involves doing some work against the attractive force. This work done is stored in the capacitor in the form of electrical potential energy. This energy may be converted to heat, light or other forms. A graph of p.d, V against charge, Q is a straight line through the origin whose gradient gives the capacitance of the capacitor.



p.d (V)

Charge, Q (C)

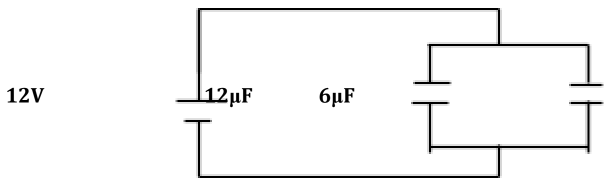
The area under this graph is equal to the work done or energy stored in the capacitor.

i.e. $E = \frac{1}{2} QV$ but $Q = CV$

Hence**$E = \frac{1}{2} CV^2 = Q^2 /2C$**

Example 9.3

1. The figure below shows two capacitors connected to a 12V supply



Determine: a) the effective capacitance of the circuit

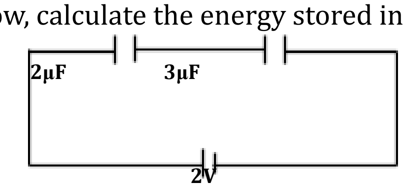
b) Charge on each capacitor

c) Energy stored in the combination

{ans. 18μF, 72μC, $2.196 * 10^{-3}$J}

a) 12+6 = 18μF b) Q1= 12*12 =144μC c) E= ½ CV^2 =1/2 *18*10^{-6}*122 =2.196*10^{-3}J

2. In the figure below, calculate the energy stored in the combined capacitor.



{ ans. $2.4*10^{-6}$)

C = 2*3 /2+3 =1.2μF

E = ½ *1.2 *10-6 *2^2 = 2.4 * 10^{-6} J

9.10.4: Application of capacitors

a) Rectification (smoothing circuits)

In the conversion of alternating current to direct current using diodes, a capacitor is used to maintain a high d.c. voltage. This is called smoothing or rectification.

b) Reduction of sparking in the induction coil

A capacitor is included in the primary circuit of the induction coil to reduce sparking.

c) **In tuning circuits**

A variable capacitor is connected in parallel to an inductor in the tuning circuit of a radio receiver. When the capacitance of the variable capacitor is varied , the electrical oscillations between the capacitor and the inductor changes. If the frequency of oscillations is equal to the frequency of the radio signal at the aerial of the radio, that signal is received.

d) **In delay circuits**

Capacitors are used in delay circuits designed to give intermittent flow of current in car indicators.

e) **In camera flash**

A capacitor in the flash circuit of a camera is charged by the cell in the circuit. When in use, the capacitor discharges instantly to flash.

TOPIC 10: CURRENT ELECTRICITY

10.1: Introduction

Electric current is defined as the rate of flow of charges.

i.e. current, I= amount of charge Q/time of flow, t

The SI unit of electric current is the ampere (A). Other smaller units include milliampere (mA) and microampere(μA):

1A= 10^3mA

1A= 10^6μA

Charge is usually measured in units called coulomb(C). When a switch in an electrical circuit is open the circuit is referred to as an open circuit and when it is closed such that current flows, it is said to be closed. So current only flows in a closed circuit. Though electrons flow from the negative terminal to the positive terminal of a cell, the conventional current direction is from positive to negative terminal of a cell.

The charge of an electron is **e** coulomb. If **n** electrons pass through a point in a circuit, the total charge Q crossing that point is given by:

Q= ne

There fore I= Q/t = ne/t

Generally, the charge of an electron is $-1.6*10^{-19}$C. In calculations, the negative sign is always ignored.

Example 10.1

1. If the current in a circuit is 2A, calculate:

a) The charge that crosses a point in the circuit in 0.6s.

I=Q/t

Q=2*0.6 =1.2C

b) The number of electrons crossing the point per second. Take e=$1.6*10^{-19}$C.

I=Q/t = ne/t

1.2/0.6 = (n*$1.6*10^{-19}$)/0.6

n= 1.2/ ($1.6*10^{-19}$) = $7.5*10^{18}$ electrons.

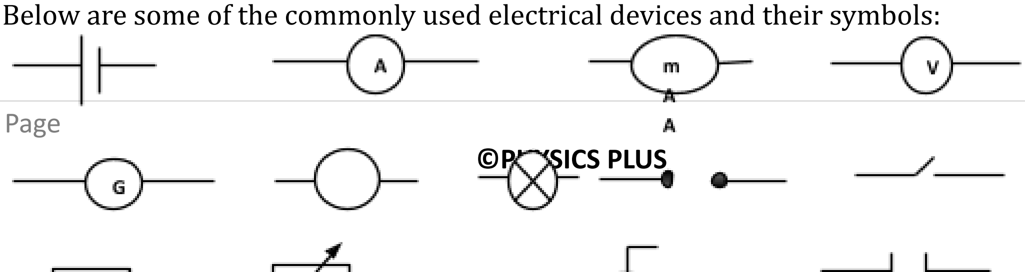
2. A charge of 180C flows through a conductor for 3 minutes. Calculate the current flowing through the conductor.

I=Q/t =180/ (3*60)

= 1A

10.2: Common electrical symbols

Below are some of the commonly used electrical devices and their symbols:



+ -

A cell **Ammeter** **Milliammeter** **Voltmeter**

OR

A galvanometer **A bulb/ filament** **a.c power supplyA witch**

A fixed resistor **A variable resistor** **A rheostat** **A capacitor**

OR

A fuseWires crossing with connectionWires crossing with no connection

10.3: Electromotive force (emf), potential difference (terminal voltage) of a cell and internal resistance

For charges to flow through a conductor, work must be done to overcome the resistance offered by the conductor. Some work is also done to drive charges through the cell itself. The energy required to do this work must be supplied by the cell itself. The maximum energy available per coulomb between the terminals of a cell when there is no resistance i.e. in an open circuit is referred to as its **electromotive force** (emf).

On the other hand, the potential difference between the terminals of a cell is the energy per coulomb between its terminals when supplying current i.e. in a closed circuit. It is also called **terminal voltage**.

V_1= the electromotive force of the cell

V_2= potential difference/ terminal voltage

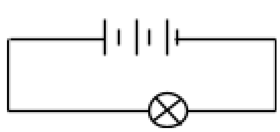
Generally, emfof a cell is larger than the terminal voltage.

The difference between the emf and terminal voltage is due to the resistance offered by the cell itself. This resistance by the cell is called **internal resistance**. The work done per coulomb to overcome internal resistance is known as the **lost volts**.

i.e. **lost volts= emf-terminal voltage**.

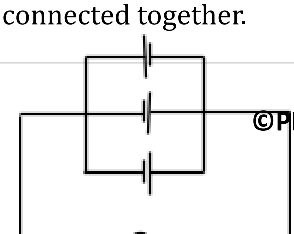
10.4: Arrangement of cells

- Suppose three cells each of emf 1.5V are connected in series, then the total emf of the circuit is the sum of the emf of the three cells.



In series arrangement of cells, a positive terminal of one cell is connected to the negative terminal of the next cell.The current flowing through the circuit will be higher and hence the bulb would be brighter than when it would have been a single cell.

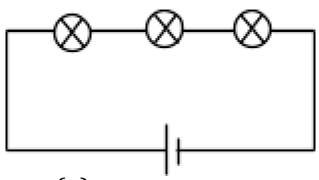
- In parallel connection of cells, all the positive terminals are connected together and all the negative terminals also connected together.



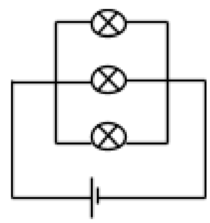
In this case, the bulb uses an emf equivalent to the emf of one cell. The current flowing in the circuit will also be lower. The advantage this method of connection has over series connection is that it can supply current for a longer time.

10.5: Arrangement of electrical components

Electrical devices can also be connected in series or parallel or even a combination of the two. Consider three bulbs connected as shown below:



(a)



(b)

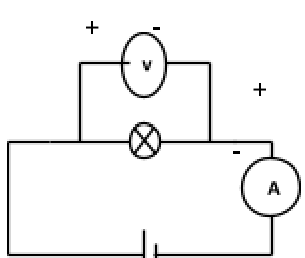
In (a), the bulbs have been connected in series. In this case, the current flowing through the bulbs is the same and is equal to the circuit current. The sum of the voltage drop across the bulbs is equal to the total circuit voltage. When one bulb is faulty, the remaining bulbs will stop working since the circuit will be incomplete.

In (b) where the bulbs have been connected in parallel, the voltage drop across the bulbs is the same and is equal to the voltage supplied by the cell. The sum of the current through the individual bulbs is equal to the circuit current. The advantage of this method of connection is that when one of the bulbs is faulty the remaining bulbs will still be working. This method is commonly used in wiring of lighting circuits in houses.

10.6: Measurement of voltage and current

In the laboratory, voltages are measured using a voltmeter while current is measured using an ammeter or a milliammeter. The voltmeter is always connected in parallel with the components where potential difference is to be measured. They offer very high resistance to the flow of current. Thus very negligible current flows through it.

Ammeters on the other hand are connected in series with the other components where current is to be measured. This is because they have very negligible resistance and hence does not interfere with the current.



Note that the positive terminal of the voltmeter or ammeter is always connected to the positive terminal of the cell or battery.

10.7: Conductors and insulators

A **conductor** is a material that allows free flow of charges through it e.g. copper, silver, aluminium, etc. There are also those materials which do not allow flow of charges through them e.g. dry wood, plastic, rubber etc. They are generally referred to as **insulators.** Conductors are further categorized as either good or poor conductors. The good conductors are generally metals and they conduct electric current at a faster rate. This is made possible by the fact that metals have numerous free electrons moving within themselves. An example of a poor conductor is graphite. Such conductors possess less number of free electrons and hence conduct slowly. Insulators do not have free electrons at all.

There are also a group of materials whose electrical conductivity lies between that of a conductor and an insulator i.e. they conduct only when under certain conditions. Examples of such materials include silicon and germanium. They are referred to as **semiconductors**.

Some liquids (solutions) are also very good electrical conductors like dilute sulphuric acid, sodium chloride and potassium hydroxide.

10.8: Sources of electricity

Below are some of the common sources of electricity:

- Chemical cells
- Generators
- Solar cells

In this topic, we are going to look at chemical cells.

1. **Chemical cells**

These cells rely on chemical reactions to produce electromotive force. Chemical cells are of two types:

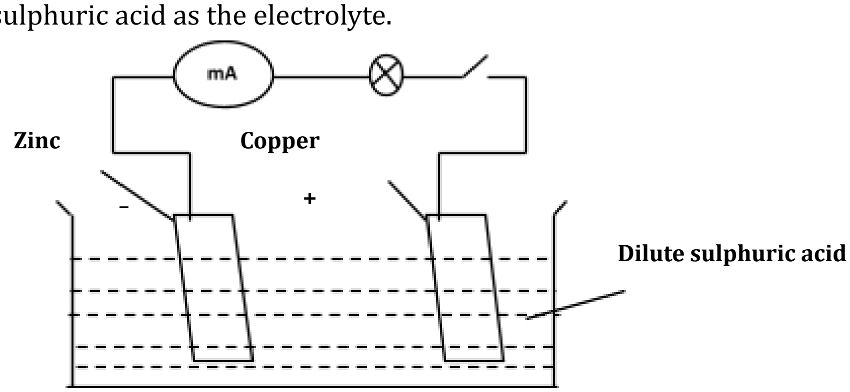
- Primary cells
- Secondary cells.

❖ **Primary cells**

These are further grouped into a simple cell and a dry cell.

a) **A simple cell**

A simple cell consists of two plates; zinc plate which is the negative plate and copper which is the positive plate and dilute sulphuric acid as the electrolyte.



With the switch open, bubbles can be observed around the zinc plate indicating that the reaction between zinc and the acid is faster than that between copper and the acid. When the switch is closed the milliammeter deflects and the bulb lights, a sign that current is flowing in the circuit.

When zinc reacts with the acid, zinc forms an ion by liberating electrons which flow through the connecting wire to the copper plate.

$Zn_{(s)} \longrightarrow Zn^{2+}_{(aq)} + 2e^-$

Meanwhile the dilute sulphuric acid dissociates into hydrogen ions (H^+) and sulphate ions (SO_4^{2-}). The hydrogen ions from the acid will move to the copper plate where they are neutralized by the electrons from the reaction between zinc and the acid. The result is formation of hydrogen gas bubbles around the copper plate.

$H^+_{(aq)} + 2e^- \longrightarrow H_{2(g)}$

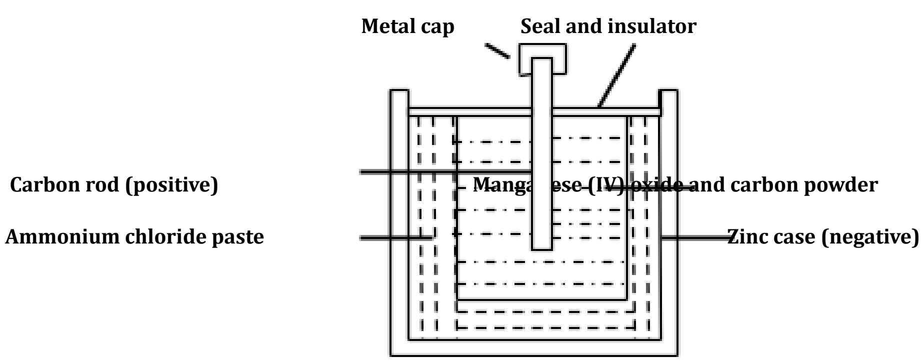
The process of formation of hydrogen gas bubbles around the copper plate is called **polarization.** This will make it difficult for electrons to flow and hence the size of the current goes down and the bulb becomes dim. Polarization can be minimized by adding a depolarizer e.g. potassium dichromate. The depolarizer should be one which does not react with the electrolyte.

Also as the zinc reacts with the acid, it dissolves and exposes the hidden impurities of carbon and iron. These impurities promote the reaction between zinc and the acid. The zinc is therefore eaten up even when there is no current being supplied i.e. in an open circuit. This is called **local action**. It can be minimized by using **pure zinc** or applying a layer of mercury on the zinc plate. The process is called **amalgamation**.

Polarization and local are referred to as cell defects.

b) Dry cell

It is called leclanché cell. The electrolyte has been replaced with ammonium chloride paste while the electrodes are now **carbon rod** as the positive terminal and **zinc casing** as the negative terminal. The carbon rod is surrounded by manganese (IV) oxide mixed with carbon powder.



When the cell is working, zinc is converted to zinc chloride liberating hydrogen ions. The hydrogen ions are neutralized by the electrons from the reaction between zinc and ammonium chloride and hydrogen gas is produced. Oxygen from manganese (IV) oxide combines with the hydrogen produced to form water. Hence polarization is minimized. However, the water formed makes the cell to be wet. Carbon powder acts as a catalyst since it is used to quicken the working of manganese (IV) oxide.

In order to minimize local action, pure zinc should be used or it should be coated using mercury. Note that local action cannot be completely eradicated.

A new dry cell has an emf of 1.5V and cannot be renewed once its energy is exhausted. A dry cell should always be stored in a dry place.

❖ **Secondary cells**

Before use a secondary cell is first charged using electricity. The energy is stored in chemical form. When the cell is in use, the stored energy is converted to electrical energy. There are two commonly used secondary cells:

- Lead-acid accumulator
- Alkaline accumulator

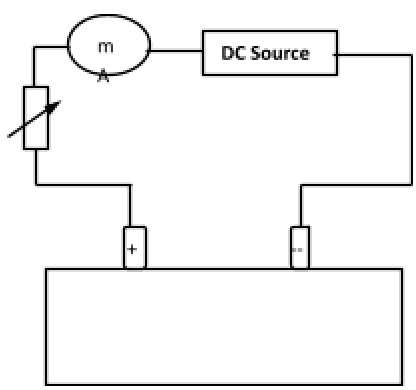
a) Lead-acid accumulator

It is the most reliable, long lasting and cost-effective secondary cell. It consists of a number of cells connected in series. Several cells connected in series forms a battery. The positive plate is lead (IV) oxide while the negative plate is spongy lead. The plates are very close to one another and are separated by insulating sheets to keep them out of contact.

The current carrying capacity of the battery depends on the surface area and the number of plates in the cell. The larger the surface area the higher the amount of charge it can store. The capacity of the cell is also directly proportional to the number of plates.

The electrolyte comprises 64% water of relative density 1.00 and 36% sulphuric acid of relative density 1.84. When the accumulator is working (discharging), water is produced which lowers the relative density of the acid. When completely discharged, the relative density of the acid is 1.18. In order to regain its energy, the accumulator is recharged using direct current. In this case, the positive terminal of the charging unit is connected to the positive terminal of the accumulator.

+ –



Recharging a lead-acid accumulator

The maximum amount of energy an accumulator can store when fully charged is referred to as its capacity. It is calculated in ampere-hour (Ah);

i.e. capacity= current in ampere * time in hours.

Note that the internal resistance of a cell is inversely proportional to the linear dimensions of the plates.

Example10.2

1. What is the capacity of a cell which can supply current of 250mA for 4hours?

Capacity= 250/1000 * 4 = 1Ah.

2. The charge stored by a cell A of plate dimensions 0.2m*0.2m is 108000C.

a) What charge is stored in cell B of plate dimensions 0.4m*0.4m?

Charge stored by A/Charge stored by B = Area of plate A/ Area of plate B

$108000C/Q_B = (0.2*0.2)/ (0.4*0.4)$

$Q_B = (108000*0.16)/0.04 = 432000C.$

b) What is the ratio of internal resistance of cell A to that of cell B?

Resistance of A α 1/ length of A

And resistance of B α 1/ length of B

Therefore, resistance of A/ resistance of B = length of B/ length of A

= 0.4m/ 0.2m =2:1

Care for the lead-acid accumulator

1. The level of the electrolyte should always be maintained above the plates. This can be done by topping up using distilled water.
2. Never draw large current from the accumulator for a longer time since this can weaken the electrodes.
3. Never leave the accumulator in a discharged condition for a long time.
4. Always keep the terminals clean.
5. Never put the accumulator directly on the ground. Instead rest it on some insulator.

c) Alkaline accumulator

This accumulator uses alkaline solution as the electrolyte e.g. potassium hydroxide. Some common types of alkaline accumulators include a nickel-iron and nickel-cadnium accumulators. These accumulators are preferred where large current may be required for emergency e.g. in hospitals.

Advantages of alkaline accumulator over lead-acid accumulator

1. Large current can be drawn from them over a short time.
2. Minimal maintenance is required.
3. They are easily portable (lighter).
4. They can be kept in a discharged state for a longer time without ruining the cells.

Disadvantages of alkaline accumulator over lead-acid accumulator

1. Their initial cost is very high i.e. expensive.
2. They have a lower emf per cell.

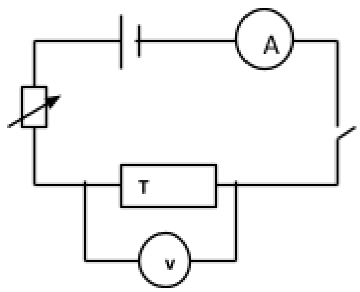
10.9: How to use an ammeter and voltmeter

- Connect the positive terminal of the ammeter/ voltmeter to the positive terminal of the battery.

- Ensure that the pointer is initially at zero i.e. there is no zero error. If there is a zero error, correct it before using the instrument.
- Select an appropriate scale to use.
- Avoid parallax error taking readings i.e. view the scale normally.

10.10: Ohm's law (form three work)

This law relates the current flowing through a conductor and the voltage drop across that section of the conductor. The law states: **the current flowing through a conductor is directly proportional to the potential difference across its ends provided temperature and other physical factors are kept constant**. The following set up can be used to investigate Ohm's law:

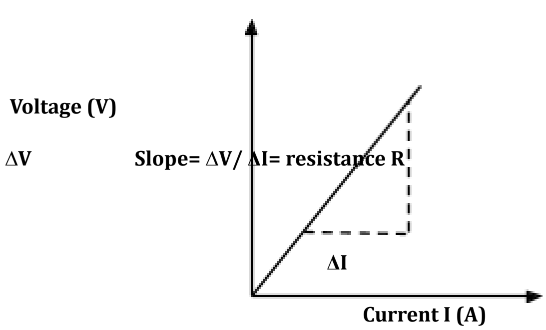


- Close the switch and adjust the current flowing through the conductor T using the rheostat to the least possible value. Record the corresponding voltmeter reading.
- Increase the current in steps recording the corresponding voltmeter readings. Record your values in the table below:

Current I (A)					
Voltage V (V)					

- Plot a graph of voltage against current. Hence determine the slope of the graph.

A graph of voltage against current is a straight line through the origin. Hence voltage drop across the conductor is directly proportional to the current through it;



$V \alpha I$

V/I = constant

The constant is known as resistance R of the conductor T under investigation.

Thus, $V/I = R$

Or $V = IR$.

Hence the slope of a voltage—current graph is equal to the resistance R of the conductor T. electrical resistance can be defined as the opposition offered by a conductor to the flow of electric current. It is measured using an ohmmeter.

The SI Unit of electrical resistance is the ohm (Ω). Other units include kilo-ohm (kΩ) and mega-ohm (MΩ);

$1\Omega = 10^{-3} k\Omega$

$1\Omega = 10^{-6} M\Omega$

Materials which obey Ohm's law are said to be ohmic materials while those which do not obey the law are said to be non-ohmic materials. The graph of voltage against current for non-ohmic materials is a curve or may be a straight line but does not pass through the origin.

The inverse on resistance is called conductance;

Conductance= 1/ resistance R.

Example 10.3

1. Calculate the current flowing through a 8Ω device when it is connected to a 12V supply.

   $I = V/R$

   $I = 12V/8\Omega = 1.5A$

10.10.1: Factors affecting the resistance of a conductor

There are three main factors that affect the resistance of a conductor:

a) **Temperature**

Increase in temperature enhances the vibration of the atoms and thus higher resistance to the flow of current.

b) **Length of the conductor L**

The resistance of a uniform conductor increases with increase in length.

c) **Cross section area A**

A conductor having a wider cross section area has more free electrons per unit length compared to a thin one. Hence a thicker material has a better conductivity than a thinner one. Generally, resistance varies inversely as the cross section area of the material.

Therefore, at a constant temperature resistance varies directly as the length and inversely as the cross section area of the conductor;

$R \alpha L/A$

R= (A constant * L/A)

Or simply, AR/L= constant

The constant is called the resistivity of the material;

Resistivity ϱ= (cross section area A * resistance R) / length L.

Resistivity is measured in ohm-metre (Ωm).

Example 10.4

1. A wire of resistance 3.5Ω has a length of 0.5m and cross section area $8.2 * 10^{-8}m^2$. Determine its resistivity.

   Resistivity $\varrho = AR/L = (8.2*10^{-8}m^2*3.5\Omega)/0.5m$

   $= 5.74*10^{-7}\Omega m$

2. Two conductors A and B are such that the cross section area of A is twice that of B and the length of B is twice that of A. If the two are made from the same material, determine the ratio of the resistance of A to that of B.

   $R=\varrho L/A$

   Therefore, $R_A= \varrho_A L_A/A_A$

   And $R_B= \varrho_B L_B/A_B$

   Where $L_B=2L_A$

   $A_B= 1/2A_A$

   And $\varrho_A= \varrho_B$

   Hence $R_A=\varrho_A L_A/A_A$ and

   $R_B= 2\varrho_A L_A/0.5A_A = 4\varrho_A L_A/A_A$

   Thus $R_A/R_B=\dfrac{\varrho_A L_A/A_A}{4\varrho_A L_A/A_A} = 1/4$

   $R_A: R_B= 1:4$

10.11: Resistors

A resistor is a specially designed conductor that offers a particular resistance to the flow of electric current. There are three main groups of resistors:

a) Fixed resistors- offer fixed values of resistance. They have colour bands around them.

b) Variable resistors- offer varying resistance e.g rheostat and potentiometer.

c) Non-linear resistors- the current flowing through these resistors does not change linearly with the voltage applied. Examples include a thermistor and light-dependent resistor (LDR).

10.11.1: Measurement of resistance

Three methods may be used:

a) Voltmeter- ammeter method

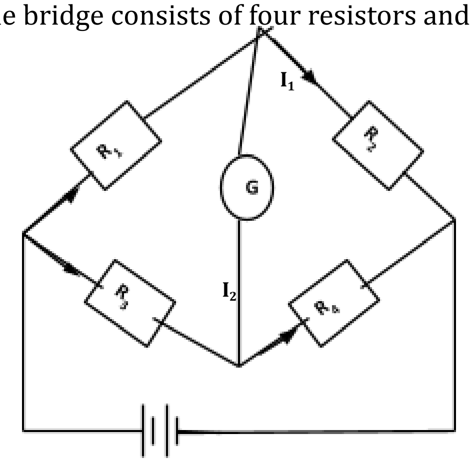
In this method, the current flowing through the material and voltage across its ends are measured and a graph of voltage against current plotted. The slope of the graph gives the resistance offered by the material.

b) The wheatstone bridge method

A wheatstone bridge consists of four resistors and a galvanometer connected as shown below:

I_1

I_2



The values of three out of the four resistors must be known. The value of one of the resistors is adjusted to a point that the galvanometer does not deflect. At this point, the voltage drop across R_1 is equal to that across R_3. Similarly, the voltage drop across R_2 is equal to that across R_4. Note that the current flowing through R_1 is equal that through R_2. Also, the current through R_3 is the same to that through R_4.

Therefore, $I_1 R_1 = I_2 R_3$…………………………… i

$I_1 R_2 = I_2 R_4$…………………………… ii

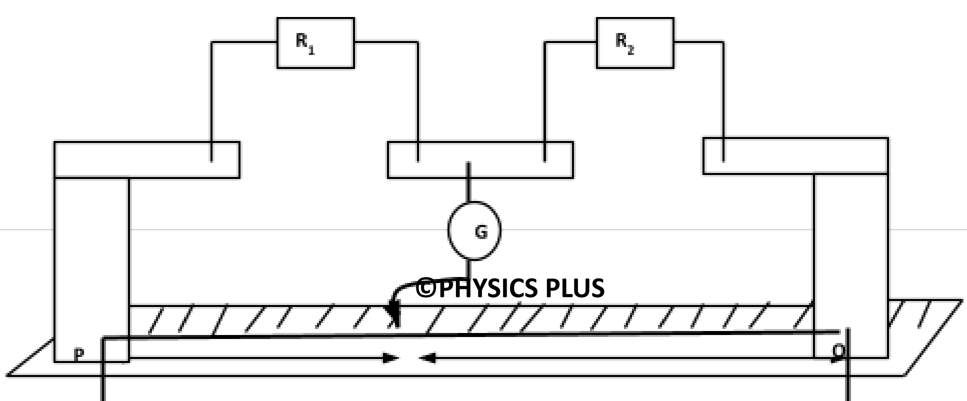
Dividing equation (i) by (ii), we get;

$R_1/R_2 = R_3/R_4$

This method is more accurate compared to the voltmeter- ammeter method since the voltmeter has some resistance against the flow of current and thus takes up some voltage.

c) The metre bridge method

This method relies on the fact that resistance is directly proportional to the length of the conductor.



L_1 K L_2

The values of R_1 and R_2 must be known. Suppose at point K the galvanometer does not deflect, then the voltage drop across R_1 equal the voltage drop across the section L_1. Similarly, the voltage drop across R_2 equals the voltage drop across the section L_2. If the current through R_1 and R_2 is I_1 and that through the section L_1 and L_2 is I_2, then;

$I_1R_1 = I_2L_1$ ……………………………….. i

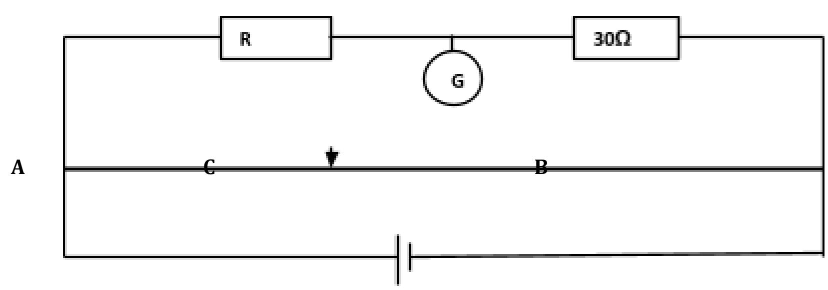
$I_1R_2 = I_2L_2$ ………………………………… ii

Dividing equation (i) by (ii), we get;

$R_1/R_2 = L_1/L_2$

Example 10.5

1. In an experiment to determine the resistance of a nichrome wire using the metre bridge, the balance point was found to be at the 40cm mark. Given that the value of the resistor to the right is 30Ω, calculate the value of the unknown resistor R.



$L_{AC}/L_{CB} = R/30\Omega$

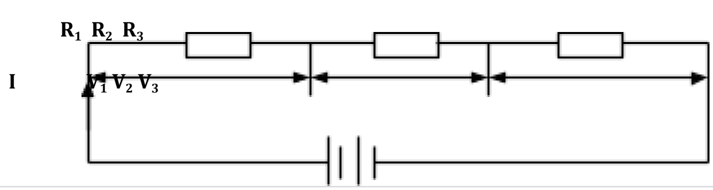
$40cm/60cm = R/30\Omega$

$R = (30*40)/60 = 20\Omega$

10.11.2: Resistor networks

a) Series network

When resistors are arranged in series the same current pass through each one of them. Consider three resistors connected as shown below:



V

From Ohm's law, V= IR.

The voltage drop across R_1; V_1= IR_1

The voltage drop across R_2; V_2 =IR_2

The voltage drop across R_3; V_3=IR_3

And the total circuit voltage V= $V_1+V_2+V_3$.

Thus V= $IR_1+IR_2+IR_3=I(R_1+R_2+R_3)$

V/I =$(R_1+R_2+R_3)$

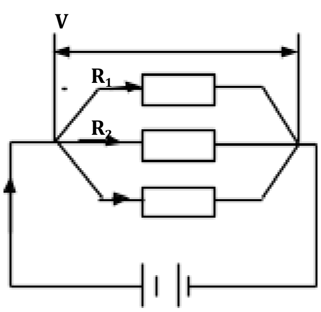
But V/I = R

Thus the combined circuit resistance R=$R_1+R_2+R_3$.

Generally, the effective resistance of resistors arranged in series is equal to the sum of the individual resistances.

b) Parallel network

When resistors are connected in parallel, the same voltage is dropped across them. Consider three resistors connected as shown below:



I R3

V

Suppose the current flowing through R_1 is I_1, through R_2 is I_2 and through R_3 is I_3 then:

The voltage drop across R_1; $V_1=I_1R_1$

The voltage drop across R_2; $V_2=I_2R_2$

The voltage drop across R_3; $V_3=I_3R_3$

But $V_1= V_2= V_3=$ V and $I=I_1+I_2+I_3$

Therefore, $I=V/R_1 + V/R_2 + V/R_3$

$I/V = (1/R_1 + 1/R_2 + 1/R_3)$

But I/V= 1/R.

Hence $1/R= 1/R_1 + 1/R_2 + 1/R_3$

R is the combined circuit resistance.

Special case of two resistors in parallel

It follows that $1/R= 1/R_1+1/R_2$

$1/R= (R_1+R_2)/R_1R_2$

Hence the effective resistance $R= R_1R_2/ (R_1+R_2)$.

Generally for n resistors arranged in parallel, the effective resistance of the arrangement is given by; $1/R=1/R_1+1/R_2+\ldots\ldots\ldots\ldots+1/R_n$

NOTE: when a circuit comprise of both series and parallel connections, the arrangement is systematically reduced to a single resistor.

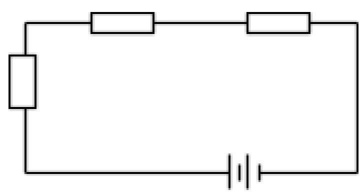
Example 10.6

1. The figure below shows 3 resistors.

5Ω3Ω

8Ω

12V



Calculate:

a) The effective resistance of the circuit.

$R= (8+5+3)\Omega = 16\Omega$

b) The total current in the circuit.

$I=V/R = 12V/16\Omega = 0.75A$

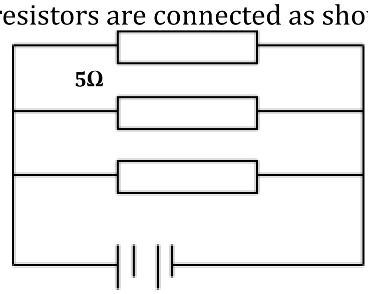
c) The voltage drop across each resistor.

$V_{8\Omega}= 0.75*8 = 6.0V$

$V_{5\Omega}= 0.75*5 =3.75V$

$V_{3\Omega}=0.75*3 =2.25V$

2. Three resistors are connected as shown below:



3Ω

6Ω

12V

Calculate:

a) The total resistance of the circuit.

$1/R= 1/5+1/3+1/6$

$1/R= (6+10+5)/30 =21/30$

$R= 30/21 = 1.4286\Omega$

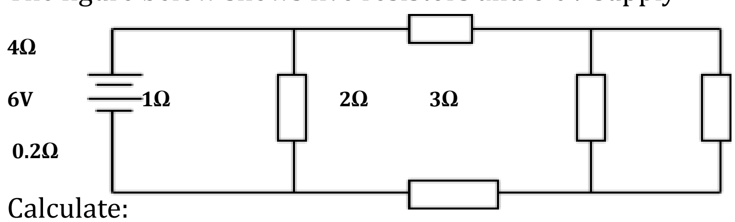
b) The current through each resistor.

$I_{5\Omega}=12V/5\Omega=2.4A$

$I_{3\Omega}=12V/3\Omega=4.0A$

$I_{6\Omega}=12V/6\Omega=2.0A$

3. The figure below shows five resistors and 6.0V supply.



Calculate:

a) The effective resistance of the circuit.

$R_{2,3\Omega}=(2*3)/(2+3) = 1.2\Omega$

$R_{4,1.2,0.2\Omega}=4+1.2+0.2 =5.4\Omega$

$R= R_{1,5.4\Omega}=(1*5.4)/(1+5.4) = 0.8438\Omega$

b) The total circuit current.

$I=V/R =6V/0.8438\Omega =7.1107A$

10.11.3: Internal resistance r

When a cell supplies current in a circuit, the potential difference between its terminals is observed to be lower than its electromotive force (emf). This difference is due to the internal resistance of the cell. Some work must be done to overcome this resistance and so the drop in the emf of the cell is responsible for this. The difference is referred to as the **lost volt** and is given by **Ir.**

i.e. lost volts= emf-terminal voltage

Or simply **emf= terminal voltage + lost volts**

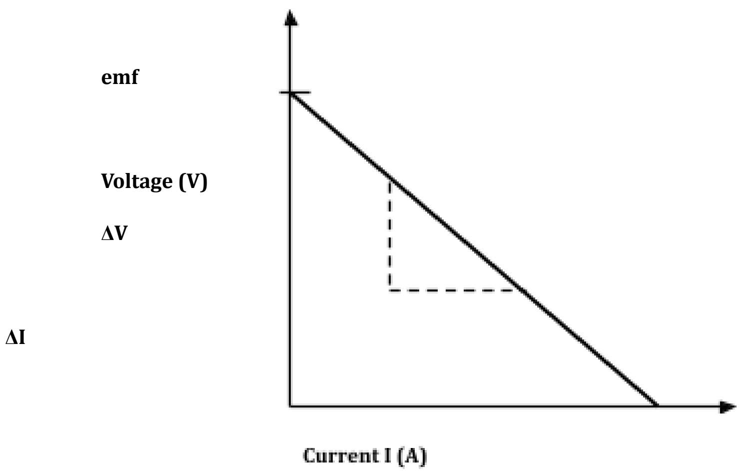
The mathematical equation connecting emf, circuit current, external resistance and internal resistance of the cell is given by:

E= IR + Ir= I(R+r).

Internal resistance of a cell can be obtained experimentally. In such an experiment, the following data was obtained:

Current I(A)	**0.1**	**0.2**	**0.3**	**0.4**	**0.6**	**0.8**
Voltage V(V)	**1.43**	**1.30**	**1.19**	**1.09**	**0.82**	**0.58**

When a graph of Voltage V against current I is plotted, the graph will appear as shown below:



The slope of the graph= **-r (-internal resistance)** while the y-intercept= emf of the cell.

TOPIC 11: MAGNETISM

11.1: Introduction

A material is said to be magnetic if it is significantly affected (attracted or repelled) by a magnet when brought closer to it. Magnetic materials are divided into two:

Ferromagnetic materials - Magnetic materials that are strongly attracted by a magnet like nickel, cobalt and iron.

Paramagnetic materials- Are magnetic materials that are weakly attracted by a magnete.galuminium.

Magnetic materials can further be grouped as soft or hard magnetic materials.

Soft magnetic materials- these materials are easier to magnetize and also lose their magnetism fast e.gsoft iron.

Hard magnetic materials- these materials take long (or difficult) to magnetize and once magnetized they do not lose their magnetism fast e.g steel.

Materials that are not affected by magnets are referred to as non-magnetic materials e.g wood, rubber, copper, plastics etc.

All magnets have two main properties namely the attractive property and the directional property.

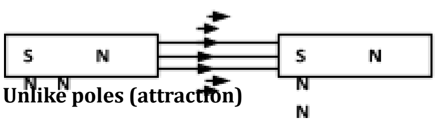
Attractive property- a magnet has two poles; the north and South pole where its attraction is strongest. This can be observed by dipping the magnet in iron filings. There will be more filings attracted around the ends.

Directional property- when a bar magnet is suspended freely in air and allowed to swing, it finally settles in the Earth's North- South direction. The end of the magnet facing the Earth's North pole is called the North- seeking pole or simply the North pole while the end facing the Earth's south pole is called the South- seeking pole or simply the South pole.

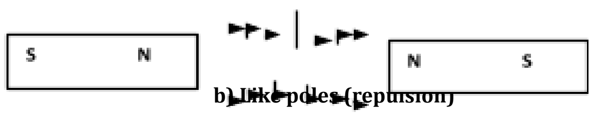
11.2: Basic law of magnetism

The law states: **like poles repel while unlike poles attract**.

A magnet will always attract a magnetic material regardless of the pole used. Also, two opposite poles will attract. However, when two similar poles are brought close together, they repel. Therefore, it is only **repulsion** that can be used to adequately identify the poles of an unknown magnet.



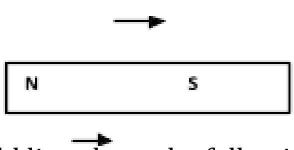
a) Unlike poles (attraction)



b) Like poles (repulsion)

11.3: Magnetic field pattern

The region around a magnet within which its influence (attraction or repulsion) is felt is known as a magnetic field. The field is stronger around the pole but weaker away from the poles. The field is represented by lines of force called magnetic field lines.



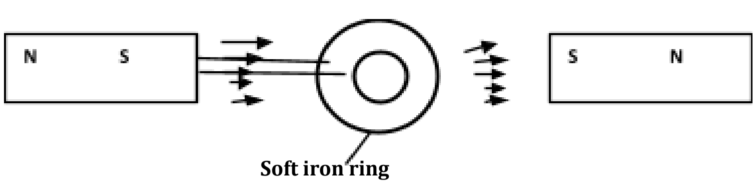
Magnetic field lines have the following properties:

- Originate from the North pole and terminate at the south pole. The direction of a magnetic field line is the direction in which a freely suspended North pole would tend to move if placed at that point in the field.
- Leave and meet a surface at right angles.
- They never cross each other.
- They are elastic in nature. They repel each other sideways.

The field lines are closer together where the field is stronger and wider apart where the field is weak. Uniformly spaced magnetic field lines imply uniform magnetic field.

11.3.1: The effect of soft iron ring in a magnetic field

When a soft iron ring or rod is placed in a magnetic field, the magnetic field lines get concentrated along the ring or rod, shielding some parts within the field from the influence of the magnet. This is called magnetic shielding or screening.



11.4: Methods of magnetization

Magnets are made through a process known as magnetization. There are four methods magnetization:

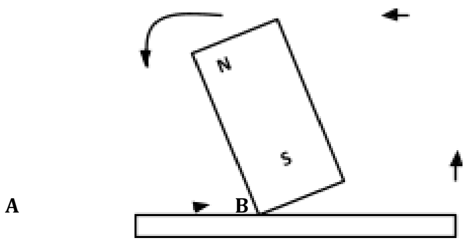
11.4.1: Induction method

When a piece of nail is allowed to hang from a magnet, it can also be used to attract another light magnetic material brought closer to it. The nail is therefore said to have been magnetized by induction. The magnetism gained in this way is temporary since when the magnet is withdrawn, the nail loses its magnetism.

11.4.2: Stroke method

This is achieved by stroking the material from one end to the other repeatedly using the same pole of a magnet. Stroking can be done in two ways:

a) Single stroke

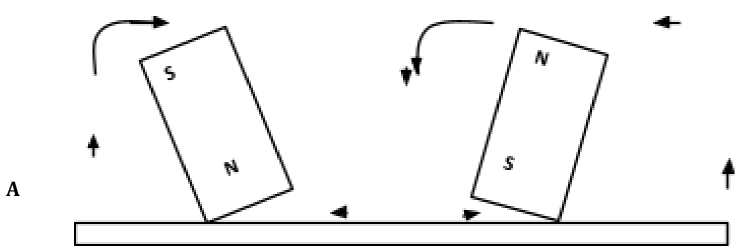


The same pole of the magnet is used to stroke the material from one end to the other repeatedly. The end of the material last touched by the magnet (end B) acquires a polarity opposite to that of the stroking pole. In this case, the stroking pole is South pole. Hence end B acquires a North pole and end A a south pole.

The magnet produced by this method has one of its poles nearer the end of the material than the other. The pole formed at the end where the stroking pole leaves the material is always closer to the end compared to that formed at the other end.

b) Double stroke

In this case the material is stroked simultaneously from the middle outwards using two magnets of opposite polarities.



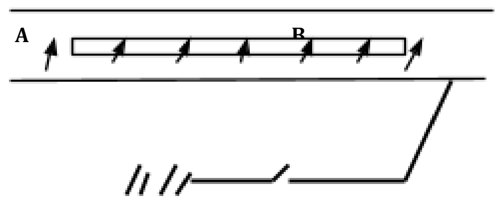
The end A acquires a South pole while B acquires a North pole. However, if the same polarities are used in double stroke method, the material acquires similar polarities at each of its ends and also similar polarities at its centre.

11.4.3: Hammering method.

It is also called the mechanical method. The material is heated until it is red hot, put to lie in the North- South direction of the earth and then hammered repeatedly. The end of the magnet facing the earth's magnetic north pole acquires a North pole and the other end a South pole. The magnet produced in this way is however weak.

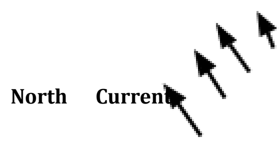
11.4.4: Electrical method

The material to be magnetized is placed inside a cylindrical coil wound with many turns of insulated copper wire. Such a coil is referred to as a **solenoid**. A direct current is then passed through the copper wire. If the current is allowed to flow around the solenoid for some time, the material becomes magnetized and can be tested for magnetism by bringing a magnetic material close to it. This method relies on the fact that current flowing through a conductor is usually accompanied with some magnetic effect as is discussed elsewhere in this book.



The polarity of the magnet produced by this method depends on the direction of flow of current. There are two rules that can be used to predict the polarity of the magnet produced:

- **The right- hand grip rule-** the rule states: if a coil carrying current is grasped in the right hand such that the fingers align in the direction of the flow of current in the coil, then the **thumb** points in the direction of the North pole.



- **The clock rule-** if on viewing one end of the material, current is flowing in the anticlockwise direction then that end will be a North pole while if it flows in the clockwise direction then that end will be a South pole.

Applying any of the above rules in our case, the end A acquires a North pole and B a South pole.

Electrical method is widely used in industries to make magnets since it is fast.

11.5: Demagnetization

Demagnetization is the process by which a magnet loses its magnetism. A magnet can lose its magnetism through the following ways:

11.5.1: Hammering

The magnet is put to lie in the East- West direction of the earth and then hit vigorously or dropped repeatedly on a hard floor.

11.5.2: Heating

The magnet is first heated until red hot and then cooled while in the East- West direction.

11.5.3: Electrical method

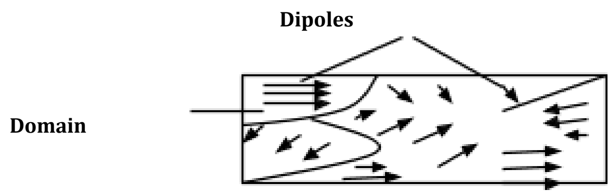
The magnet is placed inside a solenoid in the East- West direction. An alternating current is then passed around the solenoid. The alternating current reverses the arrangement of the dipoles several times per unit time thus disorienting them.

Note that during demagnetization, the magnet is placed in the East- West direction so that it does not retain some magnetism due to the Earth's magnetic field.

11.6: The Domain Theory

If a bar magnet were to be broken into smaller pieces, each piece will still have two poles; north and South pole. The smallest molecular magnet that can be obtained from a magnet is called a **dipole**.

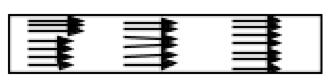
In magnetic materials (not magnets), these dipoles occupy tiny regions within the material called **domains**. The domains are separated by a thin wall. The dipoles in each domain point in a particular direction. Thus the axes of the domains in a given magnetic material point in different and random directions. The overall effect is therefore zero. Hence the material possesses no magnetism.



When the material is being magnetized, the dipoles in the domains get aligned in the same direction. The material is partially magnetized when majority of the dipoles have been aligned and it is fully magnetized when all the dipoles have been aligned. At this point, the material is said to have reached its saturation point (magnetic saturation point) and the walls between the domains disappear.

Any additional magnetizing force applied after magnetic saturation has been achieved has no effect on the magnetic strength of the material.

a) Partially magnetized



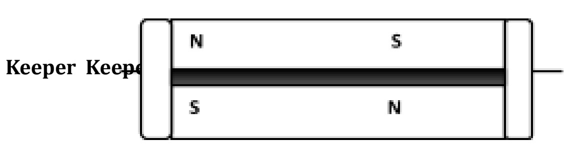
b) Fully magnetized

Demagnetization is therefore any process that interferes with the alignment of the dipoles. During demagnetization, the walls between the domains slowly return as disalignment of the dipoles occur.

11.7: Storage of magnets

All magnets lose their magnetism with time. The poles of the magnet tend to reverse the direction of the dipoles in it. This leads to **self** demagnetization. Permanent magnets can also lose their magnetism through exposure to high temperatures and mechanical shock. Therefore, there is need to take care of the magnets.

Bar magnets are stored in pairs with pieces of soft iron keepers at its ends.



The keepers get magnetized so that the dipoles of the magnet and the keeper form a closed loop.

11.8: Uses of magnets

Magnets have wide applications such as:

- Making of compasses.

- Removal of iron splinter from the eye without undergoing a surgery.

- Construction of electric bells, loud speakers, telephone receivers, transformers etc.

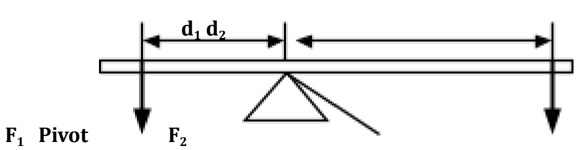
- Resetting of the index in the six’s maximum and minimum thermometer.

TOPIC 12: TURNING EFFECT OF A FORCE

12.1: Moment of a force

We have seen earlier the effects of force on a body. One such effect is force can make a body turn about a point. The turning effect of a force is referred to as **moment of that force**. it is defined as the product of the force and the perpendicular distance between the line of action of the force and the pivot;

Moment of a force =force (F)x perpendicular distance (d)



The moment of force F_1= F_1xd_1

Similarly, the moment of force F_2=F_2xd_2

The SI Unit of moment of a force is the newton-metre (Nm). The moment of a force is a vector quantity i.e has both magnitude and direction.

The knowledge of the moment of a force helps us to understand why:

- ✔ It is easier to open or close a door with the handle farthest from the hinges.
- ✔ It is easier to tighten or loosen a bolt using a longer spanner.

Example 12.1

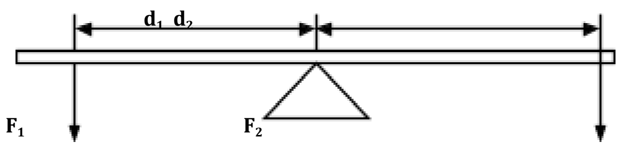
1. A force of 20N acts on a uniform rod at a distance of 25cm from the pivot. Determine the moment of the force.

Moment =force F x perpendicular distance d

=20N x 0.25m =5Nm

12.2: Principle of moments

Consider the system below:



The moment(s) to the right of the pivot is referred to as the clockwise moments while that to the left is referred to as the anticlockwise moments;

Clockwise moment= F_2d_2

Anticlockwise moment= F_1d_1

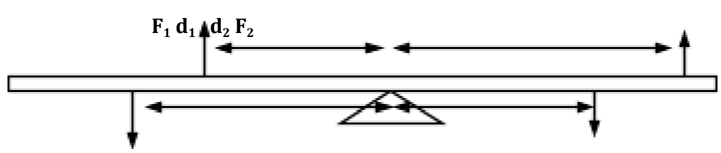
When a body is in equilibrium; **Clockwise moment= Anticlockwise moment**

The principle of moments states: **for a body in equilibrium, the sum of clockwise moments is equal to the sum of the anticlockwise moments;**

$F_2d_2= F_1d_1$

For a system of parallel forces, two conditions must be satisfied for the system to be in equilibrium:

- ✔ The sum of clockwise moments must be equal to the sum of anticlockwise moments.
- ✔ The sum of upward forces must be equal to the sum of downward forces i.e the algebraic sum of the parallel forces is zero.



d_3 d_4

F_3F_4

Clockwise moments= F_1d_1+F_4d_4

Anticlockwise moments= $F_2d_2+F_3d_3$

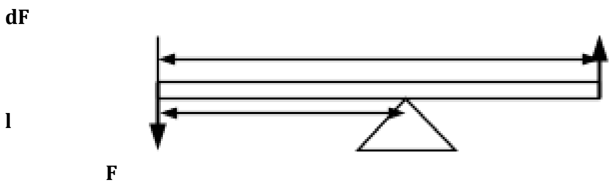
If the system is in equilibrium, then:

$F_1d_1+F_4d_4= F_2d_2+F_3d_3$

And $F_1 + F_2 = F_3 + F_4$

12.3: Moments of anti-parallel forces

Consider two equal and parallel forces acting on a body in opposite directions as shown below:



Clockwise moment= F(d-l)

Anticlockwise moment= Fl

The total moment of the system= F(d-l) + Fl =Fd-Fl+Fl = Fd

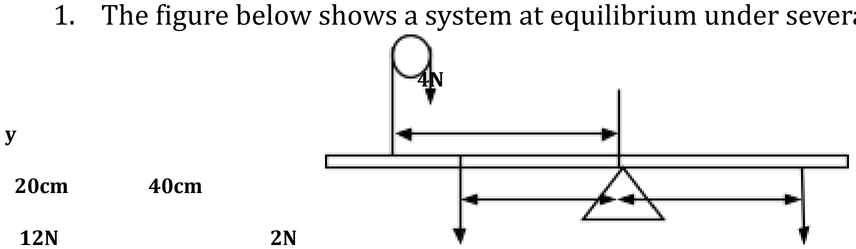
Hence the total moment of two equal anti-parallel forces is the product of one of the forces and the perpendicular distance separating them.

When two equal forces act on a body in opposite directions, one of the moments acts inwards and the other outwards. The overall effect is a turning effect on the body. Such forces are called **a couple**.

Some common situations where anti-parallel forces are applied include using the bicycle handle bars, car's steering wheel when negotiating a bend, tightening or loosening a nut, opening a water tap, movement of the bicycle pedals etc.

Example 12.2

1. The figure below shows a system at equilibrium under several forces. Find the distance y.



Clockwise moments= (2Nx40cm)+4y

Anticlockwise moment= 12Nx20cm

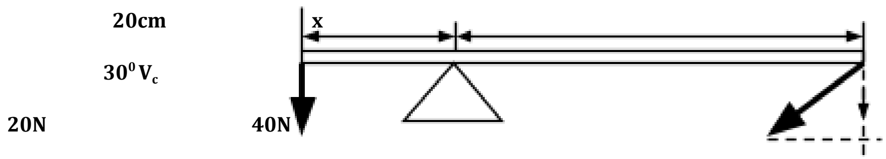
But, Clockwise moments= Anticlockwise moment

80+4y =240

4y=160

y=40cm

2. The diagram below shows a bar in equilibrium under the action of two forces.



Since the line of action of the 3N force must be perpendicular to the distance, the vertical component V_c of the force will be used;

V_c=80Cos 60 =3x0.5= 40N

Clockwise moment= 40x

Anticlockwise moment= 20N x 20cm

But, Clockwise moments= Anticlockwise moment

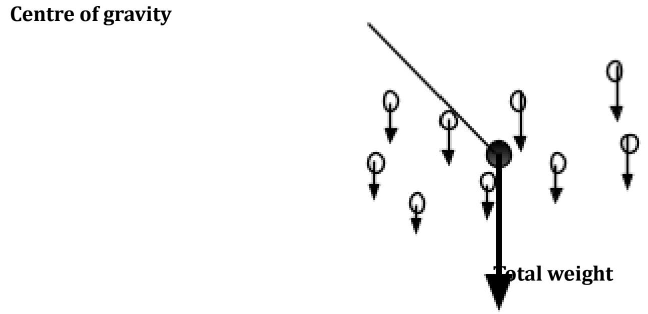
40x=20N x 20cm

x=(20x20)/40 =10cm

TOPIC 13: EQUILIBRIUM AND CENTRE OF GRAVITY

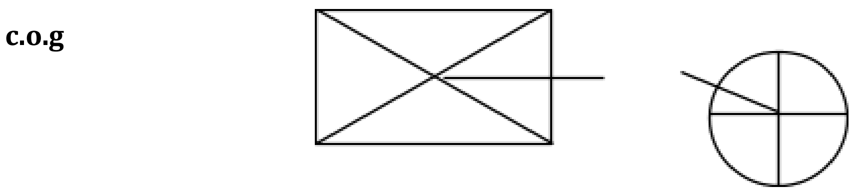
13.1: Centre of gravity

The force of gravity normally tends to attract all bodies towards the centre of the earth. The point of application of the resultant force due to the earth's gravitational pull is referred to as the **centre of gravity** of the body. This is the point at which the whole weight of the body is concentrated.



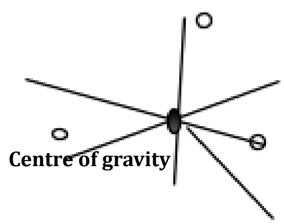
It is at this point where the body can be balanced. The centre of gravity of a body is also its centre of mass.

The centre of gravity of a regular object can be obtained through geometrical construction as it is its geometrical centre.



However, for an irregular object the centre of gravity can be located through the following steps:

- ✔ Three holes are made at different points along the edge of the object.
- ✔ The object is then suspended in turn using the holes and a plumb line attached to trace the vertical line through its centre of gravity.
- ✔ The point of intersection of the three lines will be the centre of gravity of the irregular object.



At times the centre of gravity can be in an empty space, especially for hollow objects.

13.2: Stability and equilibrium

Stability refers to the state of rest of a body while equilibrium is the state of balance of a body. When a body is in equilibrium, it stays in that position as long as there is no external influence on it.

There are three states of equilibrium namely:

- Stable equilibrium
- Unstable equilibrium

- A neutral equilibrium

These states can be illustrated by the following set up:

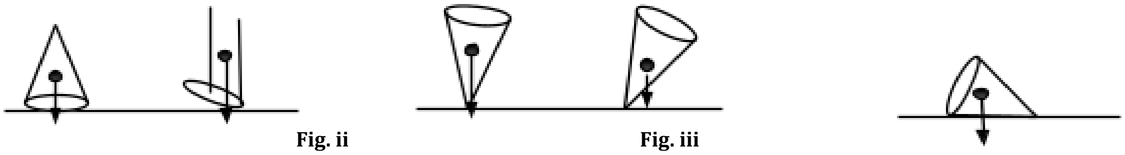


Fig. i Fig. ii Fig. iii

a) Stable equilibrium

When the cone is given a slight displacement as shown in fig. i, it is observed that the line through its centre of gravity still passes through the base area. The centre of gravity of the cone is also raised. When the force is withdrawn, the object returns to its original position without toppling. The body is said to have a stable equilibrium.

In modern buses, the luggage compartments are situated in the lower part so as to lower the position of the centre of gravity and hence increased stability.

b) Unstable equilibrium

In fig. ii, the cone has been balanced on its tip. When slightly displaced, the vertical line through its centre of gravity will fall outside its base as shown above. When the force is withdrawn, the cone will not be able to regain its original position and is thus said to be in an unstable equilibrium.

Most buses which carry heavy luggage on the upper carrier are always unstable as their centres of gravity are always raised.

c) Neutral equilibrium

In fig. iii, the cone has been placed on its side and can therefore roll over. However, this does not change the position of its centre of gravity. Hence the body is said to be in a state of neutral equilibrium.

13.2.1: Factors affecting stability

- ✔ **Area of the base**

A body with a larger base area will have the line through its centre of gravity passing through its base even when tilted through a larger angle. Hence the larger the base area the more stable a body is.

- ✔ **The position of the centre of gravity**

A body having its centre of gravity very low is more stable compared to one with a raised centre of gravity. The centre of gravity of a body can be lowered by making its base heavier.

13.3: Applications of stability

- Buses are built using lighter materials on the upper part but heavy ones on the lower part in order to lower their centres of gravity. The luggage is also put in the compartments below the passenger seats. Both these are meant to enhance stability of the bus.

- Racing cars have wider wheel bases. They are also built using heavy metals at their bases. These lower their centres of gravity thereby making them more stable and can therefore move round corners at relatively high speeds without overturning.

- A person carrying a bucket of water normally leans on the opposite side in order to adjust the position of his centre of gravity until a state of balance is achieved.

TOPIC 14: REFLECTION AT CURVED SURFACES

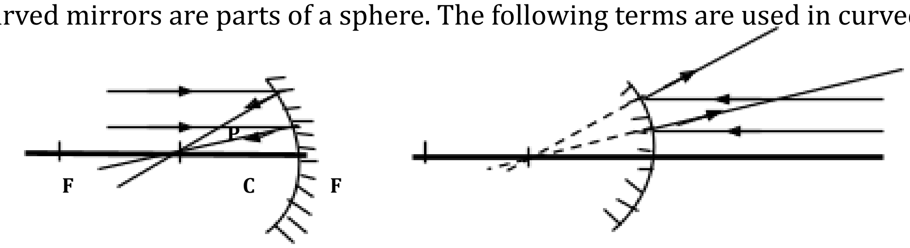
14.1: Introduction

We have already looked at reflection by plane mirrors in topic 8. When the reflecting surface is instead curved, we call it a curved mirror. There are two types of curved mirrors; concave and convex mirror. Curved mirrors whose reflecting surfaces curve inwards are called concave mirrors while those whose reflecting surfaces bulge outwards are called convex mirrors.

(a)**Concave mirror** **(b) Convex mirror**

14.2: Terms

Curved mirrors are parts of a sphere. The following terms are used in curved mirrors:



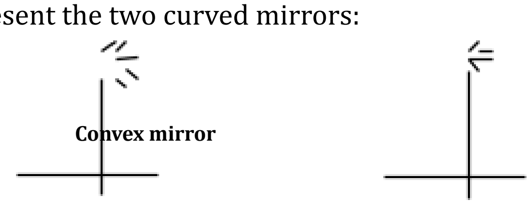
- **Pole P-** it is the centre of the mirror.
- **Centre of curvature C-** it is the centre of the sphere of which the mirror is part.
- **Radius of curvature r-** it is the radius of the sphere of which the mirror is part.
- **Principal axis-** it is a line drawn through the pole of the mirror and the centre of curvature.
- **Principal focus F** – for a concave mirror, it is the point at which all rays parallel and close to the principal axis converge at after reflection. In the case of a convex mirror, it is the point at which all rays parallel and close to the principal axis appear to diverge from after reflection.(See the figure above). It is also called the focal point.
- **Focal plane-** it is a plane perpendicular to the principal axis and passes through the focal point. It is the plane where parallel rays but not parallel to the principal axis converge at or appear to diverge from after reflection.
- **Focal length f-** it is the distance between the pole of the mirror and its focal point.

When rays are produced behind the mirror, they are indicated using dotted lines. This means that they are imaginary or virtual. Hence the focal point and focal length of a concave mirror are **real** while the focal point and focal length of a convex mirror are**virtual**. A real focal length is given a **positive sign** while a negative focal length is given a **negative sign**.

14.3: Ray diagrams

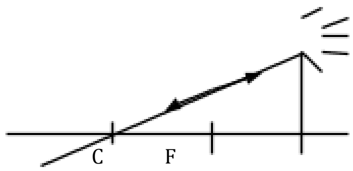
Curved mirrors form images when two rays intersect or appear to intersect. In ray diagrams, we use the following symbols to represent the two curved mirrors:

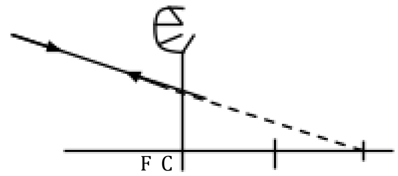
Concave mirror **Convex mirror**



There are four important rays used in ray diagrams. They include:

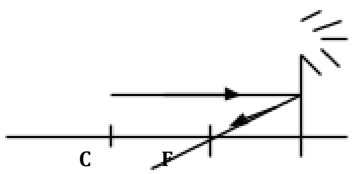
- A ray passing through C or appearing to pass through C:

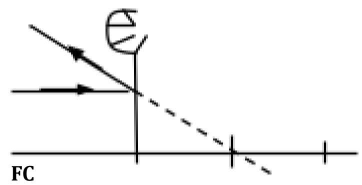




The ray is reflected along the same path.

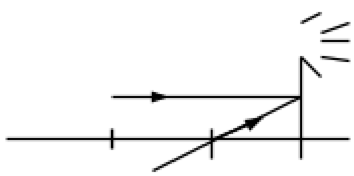
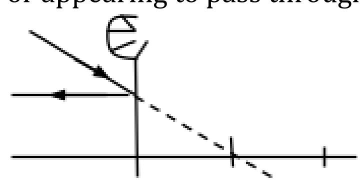
- A ray parallel and close to the principal axis.





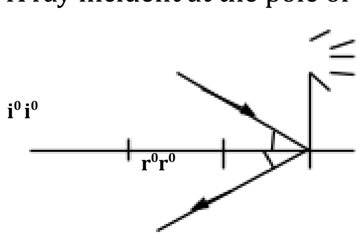
The ray is reflected through the principal focus F for a concave mirror or appear to come from the principal focus of the convex mirror.

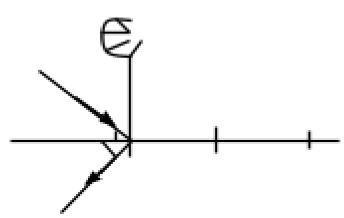
- A ray passing through the principal focus F or appearing to pass through F

The reflected ray moves parallel to the principal axis (**by the principle of reversibility of light**).

- A ray incident at the pole of the mirror.



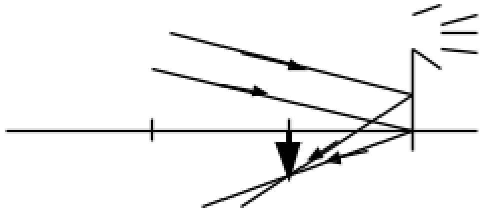
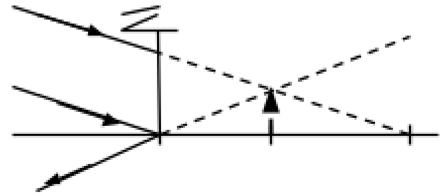
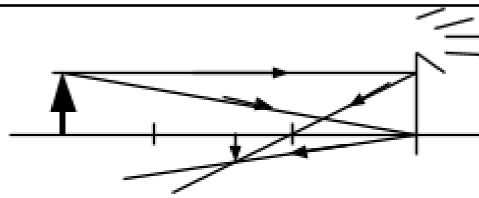
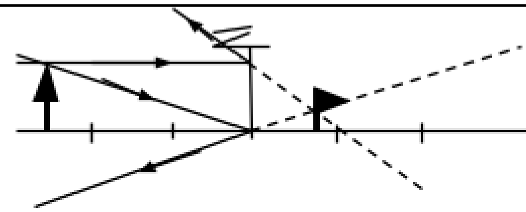
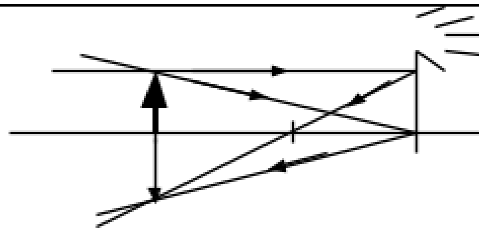
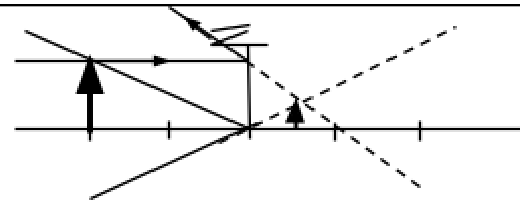
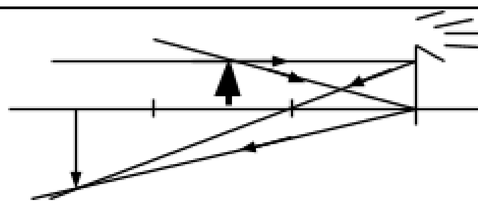
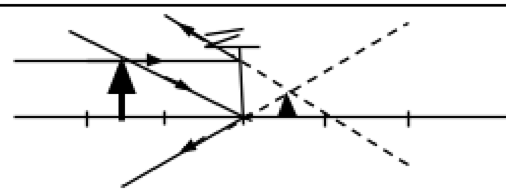
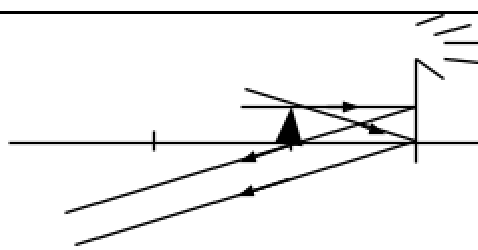
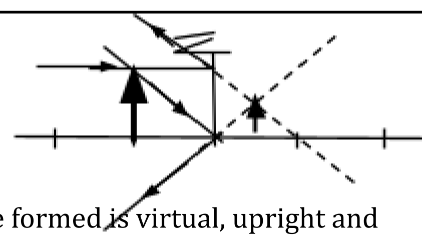


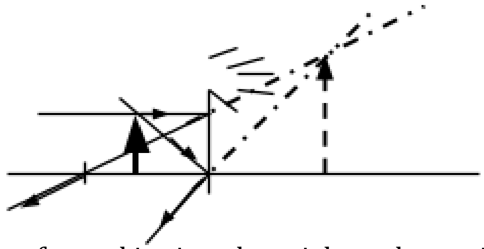
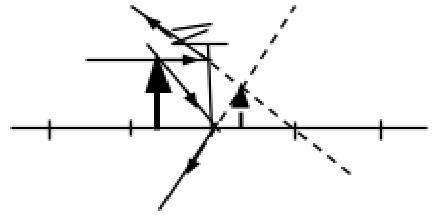
The ray is reflected making the same angle with the principal axis as the incident ray.

14.4: Image formation by spherical mirrors

The table below provides a summary of how a concave and convex mirror forms images:

Position of the object	**Image formation by concave mirror**	**Image formation by convex mirror**

Object at infinity	 Image formed is inverted, real , diminished and formed at F.	 Image formed is upright, virtual and diminished.
Object beyond C	 Image formed is real, inverted and diminished.	 Image formed is virtual, upright and diminished.
Object at C	 Image formed is real, inverted and same size as the object	 Image formed is virtual, upright and diminished.
Object between C and F	 Image formed is real, inverted and magnified.	 Image formed is virtual, upright and diminished.
Object at F	 Image formed is real, inverted and at infinity.	 Image formed is virtual, upright and diminished.

Object between F and P	 Image formed is virtual, upright and magnified.	 Image formed is virtual, upright and diminished.

Note that a concave mirror always forms real and inverted and images except when the object is placed between the focal point and the pole of the mirror when it forms a virtual and inverted image. On the other hand, a convex mirror always forms a virtual, erect and diminished image.

A real image is that image formed by actual intersection of real rays while a virtual image is formed by imaginary rays. Furthermore, a real image can be formed on a screen while a virtual image cannot be formed on a screen.

14.5: Linear magnification and the mirror formula

Linear magnification is defined as the ratio of the image size to the object size;

M= Image size/ Object size.

Similarly, it can be expressed as the ratio of the distance of the image to the distance **v** of object **u** from the mirror. Magnification has no unit.

Suppose an object is placed u cm in front of a spherical mirror of focal length f such that the image is formed v cm from the mirror, then u, v and f are related by the equation;

$1/f = 1/u + 1/v$.

This equation is referred to as the **mirror formula**. The formula holds for both concave and convex mirrors.

When applying the mirror formula, it is necessary to observe the following points:

- ✔ That all distances are measured from the mirror as the origin.
- ✔ All real distances are positive while all virtual distances are negative.
- ✔ A concave mirror has a positive focal length while a convex mirror has a negative focal length.

Example 14.1

1. A concave mirror of focal length 10cm forms a virtual image 5cm high and 30cm from the mirror. By accurate scale drawing, determine:

    a) The position of the object.

b) The height of the image.

c) The magnification of the image.

2. Determine the position, size and nature of the image of an object 4cm tall placed on the principal axis of a concave mirror of focal length 15cm at a distance 30cm from the mirror.

u=30cm, f= 15cm, h_o=4cm

1/v=1/f-1/u

= 1/15 – 1/30 =1/30

v=30cm

Also, m=v/u =h_i/h_o

Thus, h_i=(30cm x 4cm)/30cm =4cm.

Thus the image formed is real and same size as the object.

3. A convex mirror of focal length 9cm produces an image on its axis 6cm from the mirror. Determine the position of the object.

   f= -9cm, v= -6cm.

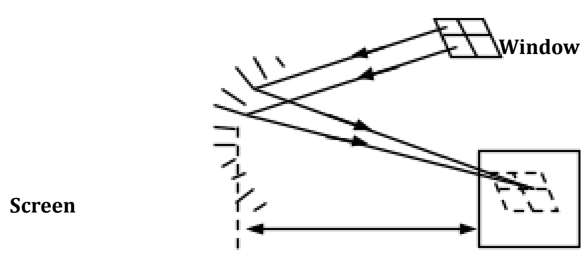
   1/u = 1/f – 1/v = -1/9 – (-1/6)

   1/u= (-2+3)/ 18 =1/18

   u=18cm

14.6: Determination of the focal length of a concave mirror

The focal length of a concave mirror can be estimated by focusing a distant object on a screen. Parallel rays from a distant object converge at the focal plane of the mirror.



f

The distance between the mirror and the screen is the estimated focal length of the concave mirror.

Alternatively, the object can be placed in front of the concave mirror at various distances and measuring the corresponding image distances then applying the mirror formula. From the values of u and v, appropriate graph is plotted. There are three possible graphs that can be obtained from the mirror formula. These include:

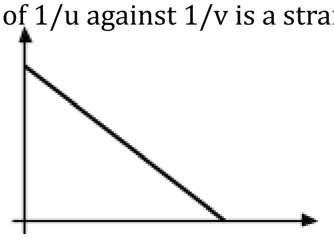
- **A graph of 1/u against 1/v**

Rearranging the mirror formula, we have;

$1/u = -1/v + 1/f$

Hence a graph of 1/u against 1/v is a straight line having a negative gradient.

1/u (cm^{-1})



1/v (cm^{-1})

At the y-intercept, 1/v=0. Substituting this in the mirror formula, we obtain;

1/f=1/u. thus the y-intercept is equal to 1/f.

Similarly, at the x-intercept 1/u=0. Substituting in the mirror formula, we obtain;

1/f=1/v i.e the x-intercept is equal to 1/f.

If the values of f obtained from the y-intercept and x-intercept above are different then we determine their average.

- **A graph of uv against u+v**

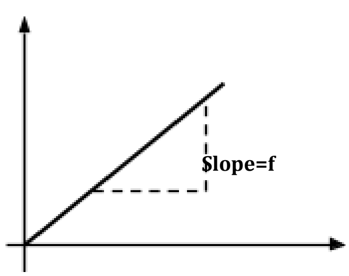
From the mirror formula, we have;

$1/f = 1/u + 1/v = (v+u)/uv$

And $uv = f(u+v)$

Hence a graph of **uv** against **u+v** is a straight line through the origin whose slope gives the focal length of the mirror.

uv (cm^2)



(u+v) cm

- **A graph of magnification M against v**

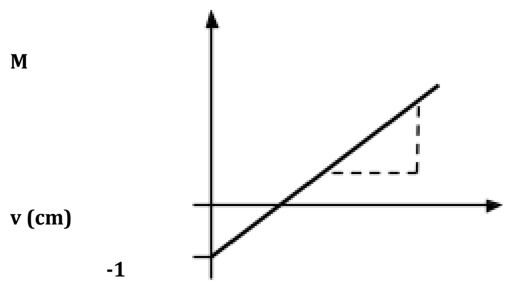
Also from the mirror formula, if we multiply through by v, we obtain:
v/f =v/u + v/v
But v/u =magnification M.

Then, v/f =M+1.

Rearranging the equation we have;

M=v/f -1

Hence a graph of magnification M against v is a straight line whose slope is equal to **1/f** and y-intercept is **-1**.



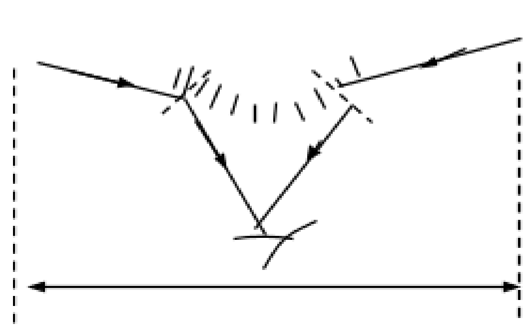
14.7: Applications of spherical mirrors and curved reflecting surfaces

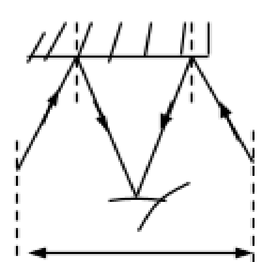
14.7.1: Concave mirrors

Concave mirrors are used by dentists to magnify teeth during extraction and as a shaving mirror. In both cases, the mirror forms upright and magnified images. Concave reflectors are also used in projector lamps, solar concentrators and telescopes.

14.7.2: Convex mirrors

Convex mirrors can be used in supermarkets to monitor customers' activities and as driving mirrors in cars and motorcycles. This is because convex mirrors provide a wide field of view compared to a plane mirror.





Wider field of view **Narrow field of view**

(a) A convex mirror **(b) A plane mirror**

However, a convex mirror used as a driving mirror has one limitation. It forms a virtual diminished image which appears to be farther away from the observer than it actually is. This may lead to misjudgment by the observer which can result into accidents.

TOPIC 15: HOOKE'S LAW

15.1: Introduction

The following are some of the notable properties of materials:

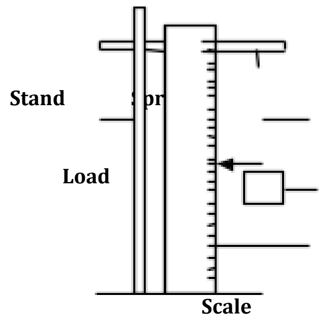
1. Strength- it is the ability of the material to breakage when subjected to a stretching or compressing or a shearing force. A strong material is one that can withstand a large force without breaking.
2. Stiffness- is the ability of a material to resist forces which tend to change its shape or size. A stiff material resists bending and is not flexible.
3. Ductility- is the ability of the material to elongate when subjected to a stretching force and undergoes plastic deformation until it breaks. Such materials are lead, copper, plasticine, etc. ductile materials can be rolled into sheets and wires.
4. Brittleness- brittle materials are fragile and do not undergo any noticeable extension on stretching but snap suddenly. Such materials include glass, dry biscuits, chalkboard chalk, etc.
5. Elasticity- is the ability of a material to recover its original shape and size after the forcing causing the deformation is withdrawn. A material which fails to recover shape but is permanently deformed is said to be plastic.

15.2: Experiments to show Hooke's law

Experiment 1

Aim: To investigate the stretching of a spiral spring

Apparatus: A spiral spring with a pointer attached**,**ametre rule**,** retort stand, two sets of clampsand20g masses.



Procedure

- Arrange the apparatus as shown above. Note the position of the pointer, P_0 without any load suspended.
- Suspend a 20g mass at the end of the spring and note the new position of the pointer, P.
- Increase the load in steps of 20g and record the corresponding pointer readings. **Precaution should be taken not to overstretch the spring.** Determine the extension, e of the spring in each case and complete the table below:
-

Mass on spring, m (g)	Stretching force, F (N)	Extension, e (m)	F/e (Nm^{-1})
0			
20			
40			
60			
80			
100			

- Unload the spring and observe what happens to the spring.
- Plot a graph of the stretching force, F against extension, e.

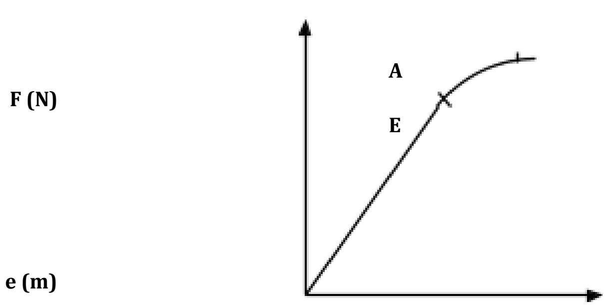
Observations

As long as the load used is not too large, the spring always returns to its original length when the force is withdrawn.

The ratio F/e is a constant.

A graph of force, F against extension, e is a straight line through the origin.

If the stretching force is increased beyond a certain value, permanent deformation occurs and the spring cannot regain its original length.



Point E is referred to as the elastic limit of the spring. Beyond the elastic limit, an increase in the stretching force leads to permanent deformation. Point A is called the yield point. Beyond this point the material breaks.

Conclusion

The above observations can be summarized in **Hooke's law** which states**: for a helical spring or any other elastic material, the extension is directly proportional to the stretching force provided the elastic limit is not exceeded.**

i.e. $F \propto e$

$F = ke$ where k is the spring constant

$k= F/e$

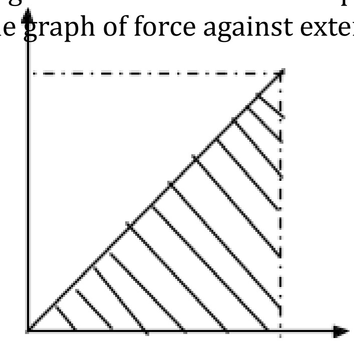
A spring constant has the unit newton per metre (N/m).

Hence the gradient of a force—extension graph equals the spring constant.

The spring constant of a spring is dependent on the material of the spring, cross-section area of the wire used, length of the spring and the number of turns per unit length.

The area under the graph of force against extension gives the work done in stretching the spring;

F **F (N)**

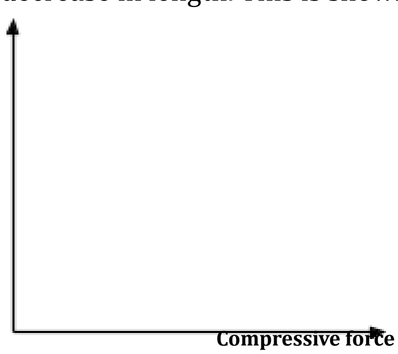


e **Extension (cm)**

Work done = ½ Fe $=1/2\ ke^2 = F^2/2k$

Alternatively, when a spring is compressed, it shortens until a point when any increase in the compressive force causes no noticeable decrease in length. This is shown on the graph below:

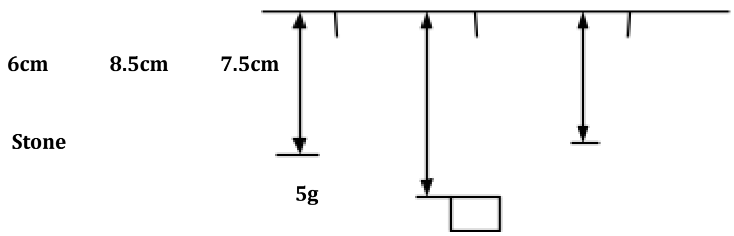
Length (m)



Example 15.1

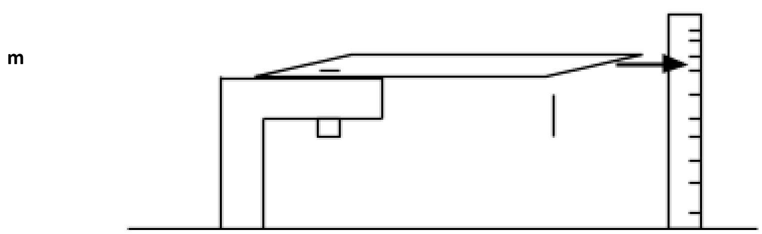
1. A spring stretches by 1.2 cm when a 600g mass is suspended on it. What is the spring constant? {ans. 500N/m}
   $K = F/e = 6N/0.012m = 500N/m$ or 5N/cm

2. The figure below shows a spring when unloaded, when supporting a mass of 5g and when supporting a stone. Use the diagram to determine the mass of the stone. {ans. 3g}



Solution
When force, F= 0.05N, e = 8.5-6.0 = 2.5cm
Therefore, k = F/e = 0.05N/2.5cm =0.02N/cm.
So F = ke
When e = 7.5-6.0 = 1.5cm, F = 0.02N/m * 1.5cm = 0.03N
Hence m = F/g =0.03/10 * 1000 = 3g

Experiment 2
Aim: To investigate the sagging of a loaded a beam.
Apparatus: 2 metre rules; one with a pointer attached at one end**,** six 100g masses**,** G- clamp**,** complete retort stand**,** a string about 30cm longanda weight holder.



Procedure
- Set up the apparatus as shown above.
- Record the initial pointer position, P_0 when no load is suspended on the beam.
- Suspend a 100g mass on the beam and record the new pointer position, P_n. Determine the amount of sagging of the beam i.e. $P_n - P_0$
- Add more loads in steps of 100g and record the new pointer position in each case. Complete the table below:

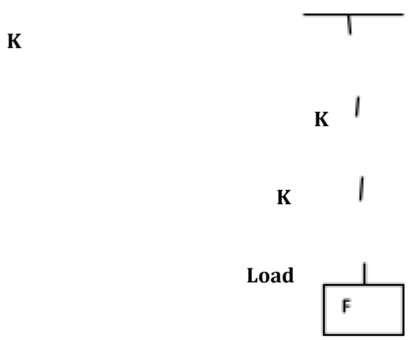
Mass, m (g)	Load, L(N)	Amount of sagging (m)
100		
200		
300		
400		
500		
600		

- Plot a graph of load, L against the amount of sagging.

15.3: Arrangement of springs

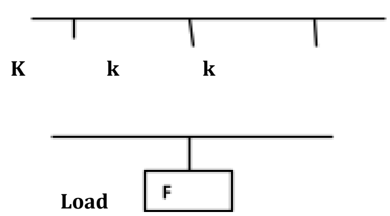
15.3.1: Series arrangement

Consider three identical springs of spring constant k.



If one spring extends by e cm when supporting a load F, then for three springs arranged in series the total extension is 3e cm when supporting the same load.
Thus for one spring, k=F/e
For the three springs, the combined spring constant $k_c = F/3e$
But F/e = k. hence $k_c = 1/3\ k$
In general, for n identical springs each of spring constant k arranged in series the combined spring constant k_c can be calculated from the equation $k_c = 1/n\ k$
If the springs have different spring constants then the combined spring constant is given by;
$\mathbf{1/k_c = 1/k_1 + 1/k_2 + 1/k_3 + \ldots + 1/k_n}$

15.3.2: Parallel arrangement



If the extension on each spring is e cm when supporting a load F, then for the three springs in parallel they share the load and thus each spring extends only by e/3 cm. hence the combined spring constant $k_c = F/(e/3) = 3F/e$.
But F/e = individual spring constant, k.
Therefore $k_c = 3k$.
Generally for n identical springs each of spring constant k, arranged in parallel, the combined spring constant k_c is given by; $k_c = nk$. If the springs are different i.e. different spring constants, then the combined spring constant k_c is given by; $\mathbf{k_c = k_1 + k_2 + k_3 + \ldots + k_n}$

TOPIC 16: WAVES

16.1: Introduction

A wave is a continuous disturbance of a medium. A pulse is short lived disturbance produced by a single vibration. Waves may be put into the following uses:

- Radio waves are used for communication e.g in radio, television, mobile phones, remote control systems, etc.
- Microwaves are used for cooking.
- Water waves are harnessed for the production of hydropower.
- Sound waves are used in ultra-sounding in hospitals.

Note that waves can also be dangerous to the environment. For instance, earthquakes which produce shock waves can lead to damage of buildings, infrastructure if its magnitude is very big.

16.2: Types of waves

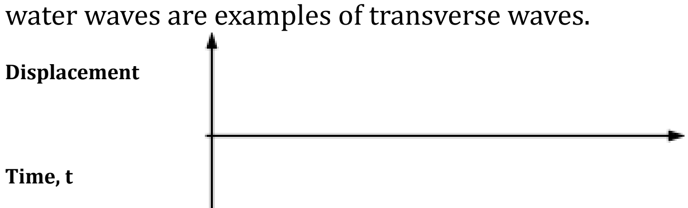
Waves are broadly categorized into two namely electromagnetic and mechanical waves.

- **Electromagnetic waves –** these waves do not require a material medium for their transmission. They include radio waves, radiant heat, light and microwaves, etc.
- **Mechanical waves –** they require material medium for their transmission. They include water waves, sound waves, etc.

Waves can be further classified in terms of the propagation of their particles as follows:

1. **Transverse waves**

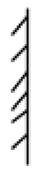
In this case, the wave motion is perpendicular to the vibration of the particles. Electromagnetic waves as well water waves are examples of transverse waves.



2. **Longitudinal waves**

This is a wave in which the wave motion is along or parallel to the direction of vibration of the particles of the medium. Sound waves are examples of longitudinal waves. When a slinky spring is stretched and released along a smooth surface with one end fixed, the spring vibrates to and fro generating regions of compression and regions of rarefaction.

Rarefaction



Compression

A region of compression is where the particles of the medium are under great pressure and are therefore closely packed together. A region of rarefaction is where the particles are spread out due to low pressure.

Some waves continuously move away from the source and are referred to as **progressive waves**.

16.3: Characteristics of waves

The following terms are associated with a wave:

a) Oscillation

It is a single complete to and fro movement from the mean position.

b) Amplitude

It is the maximum displacement of the wave particles on either side from the mean position. It is measured in metre. The position of maximum displacement is called a crest while the position of minimum displacement is called a trough.

c) Frequency, f.

It is the number of complete oscillations or cycles made in one second. It is measured in hertz (Hz). It can also be expressed as number of oscillations per second.

d) Period, T.

It is the time taken to make a complete oscillation or cycle. It is measured in second. Note that, **f=1/T or T=1/f.**

e) Wavelength, λ

This is the distance travelled by the wave in one periodic time. It can also be defined as the distance between two successive particles which are in phase e.g. between two successive crests or troughs. Two particles of a waveform are said to be in phase if they are at similar positions and are travelling in the same direction.

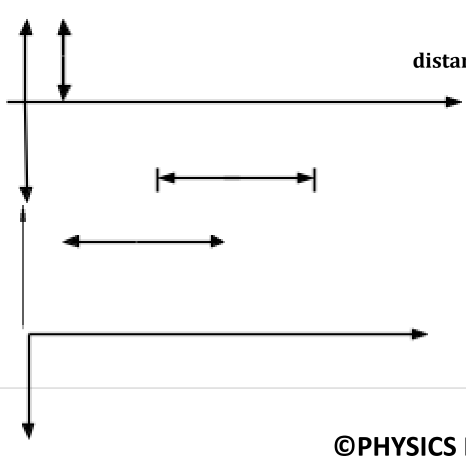
A

Displacement **distance**

λ

T

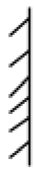
Displacement



time, t

The wavelength of a longitudinal wave can be measured as shown below:

λ

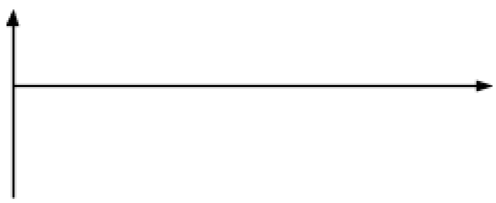


λ

f) Waves in phase.

Displacement

Distance



g) Waves out of phase (phase difference=180^0)

Displacement

Distance

Activity 16.1

1. Show two waves which are 90^0 out of phase.

16.4: The Speed of a wave

Since wavelength is the distance covered by the wave in one period, the wave speed can be determined as follows:

Speed= distance/time = wavelength, λ/period, T

$V=\lambda/T$

But $T=1/f$

Hence $v = f\lambda$.

Thus wave speed is the product of its frequency by wavelength. This formula is true for all types of waves.

Example 16.1

1. A source of waves has a frequency of 512Hz. Calculate the wavelength of the waves produced if the speed of the waves is 330m/s.

λ=v/f =330/512 = 0.64m

2. The wheel of a car is rotated at 120 revolutions per second, calculate:

a) The period, T

b) The speed of the waves generated if the wavelength is 1.5 * 10^{-2}m.

T= 1/120 = 0.0083s

V= fλ=λ/T= (1.5 * 10^{-2})/ (8.3*10^{-3}) = 1.802m/s.

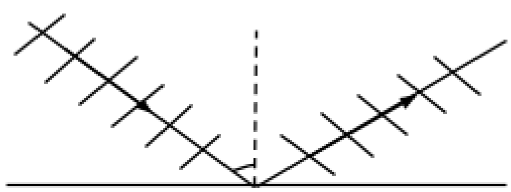
16.5: Properties of waves(form three)

Wave properties refer to the behaviour of waves under certain conditions. They include reflection, refraction, diffraction and interference among others. They can be investigated using a ripple tank which consists of a transparent tray containing water, a lamp for illumination, a white screen underneath and an electric motor (a vibrator). The motor is connected to a straight bar which produces straight waves. If circular waves are required, the bar is raised and a small spherical ball fitted to it to produce circular waves. To view the waves with ease, a stroboscope is used. A stroboscope is a disc having equally spaced slits. It is rotated and its speed controlled such that the waves appear stationary i.e frozen.

16.5.1: Reflection of waves

All waves undergo reflection. It is the bouncing back of waves when they hit an obstacle. All waves undergoing reflection obey the laws of reflection as earlier stated.

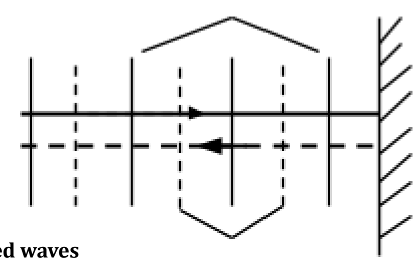
i^0 r^0



Note that the wavelength of the waves remains unchanged. The pattern of the reflected waves depends on the shape of the incident waves and the reflector. Below are some patterns:

a) Plane waves incident on a straight reflector

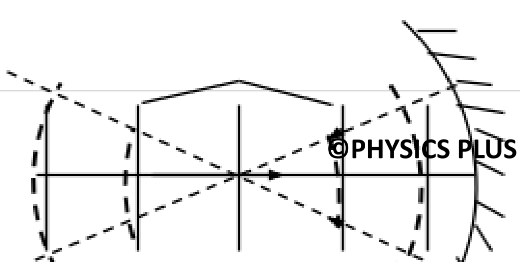
Incident wavefronts



Reflected waves

b) Plane waves incident on a concave reflector

Incident waves



F

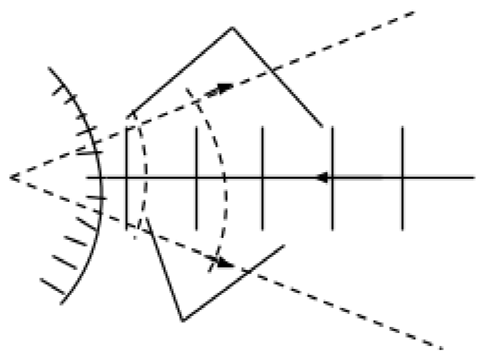
Reflected waves

The waves converge at the principal focus F after reflection.

c) Plane waves incident on a convex reflector

Incident waves

F

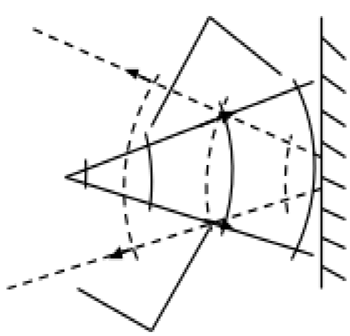


Reflected waves

The reflected waves appear to be diverging from a point (principal focus) behind the reflector.

d) Circular waves incident on a straight reflector

Incident waves



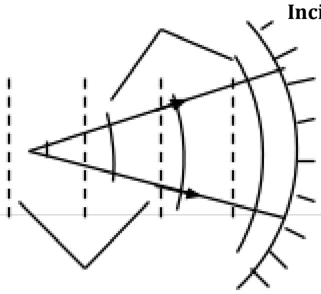
Reflected waves

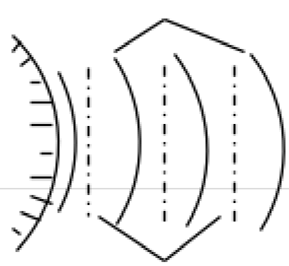
The reflected waves diverge away from the reflector.

e) Circular waves incident on a concave/convex reflector

Incident waves

Incident waves





Reflected waves

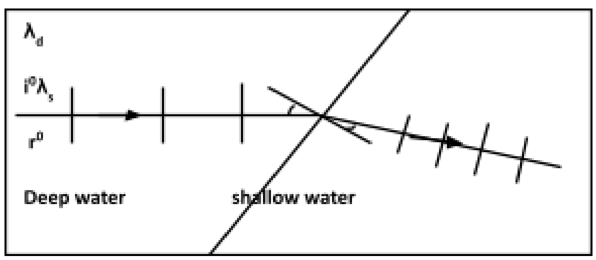
Reflected waves

16.5.2: Refraction of waves

This is the bending of waves as they travel from one medium into another. In the process, the speed of the waves changes from one medium to another. In the case of water waves, refraction occurs as the waves move from a region of a certain depth into another region of a different depth i.e. from a shallow region to a deeper region or vice versa. In general, the speed of water waves is greater in a deeper region than in a shallow region. It is important to note that the source of waves remains the same regardless of the depth thereafter. Hence, the frequency of the waves is a constant.

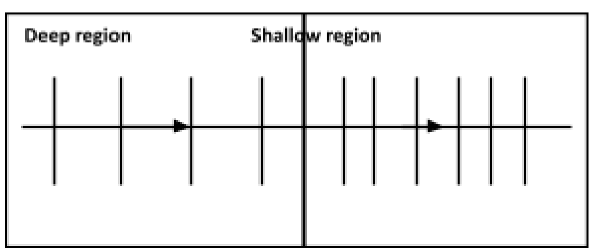
Recal**l: wave speed= frequency f*wavelength λ**.

From the equation, it is clear that when the wave speed increases the wavelength also increases and vice versa. Thus, the wavelength is longer in deeper regions than in shallow regions.



To obtain a shallow region in a ripple tank, a transparent glass block is placed in the tank with one end of its edge parallel to the vibrating bar.

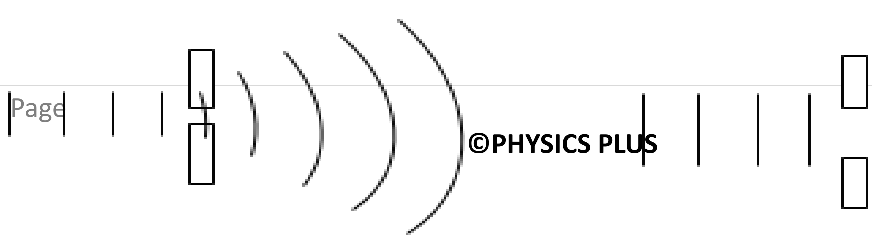
However, when the waves strike the boundary normally/perpendicularly, no bending occurs even though the speed and hence the wavelength changes.



Refraction of sound waves can be used to explain the long range of sound at night compared to daytime. This has been explained in the 'topic refraction of light'. TV and radio signals from a distant station also undergo a series of refraction and total internal reflection in the ionosphere towards the earth's surface making their reception possible.

16.5.3: Diffraction of waves

Diffraction may be defined as the spreading of waves behind an obstacle. When the aperture is nearly the same size as the wavelength of the waves, the waves emerge as circular waves spreading out around the obstacle as shown in (a) below. However, when the size of the aperture is relatively wider than the wavelength of the waves, the waves pass through as plane waves bending slightly at the edges as shown in (b).



a) Diffraction through a small aperture **b) Diffraction through a wide aperture**

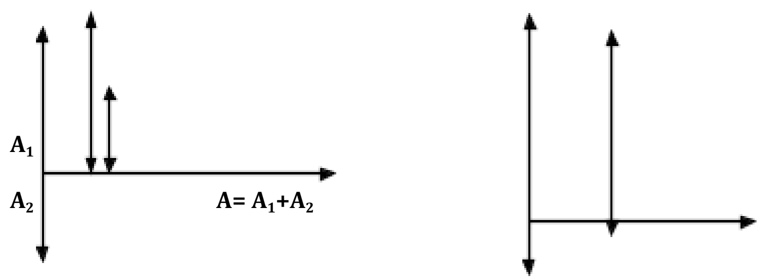
Diffraction of sound waves can be used to explain why sound within a room can be heard round a corner without necessarily having to see the source of the sound.

Diffraction of light waves is not a common occurrence due to their shorter wavelengths. Nevertheless, diffraction of light waves can be observed when light pass through a small opening at the roof of a dark room. A shadow which is broader than the opening forms on the floor of the room.

16.5.4: Interference of waves

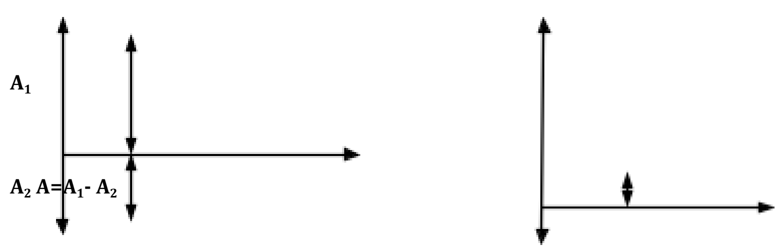
Interference occurs when two waves merge. Such a merger may give rise to three cases:

- A much larger wave is formed i.e. constructive interference.



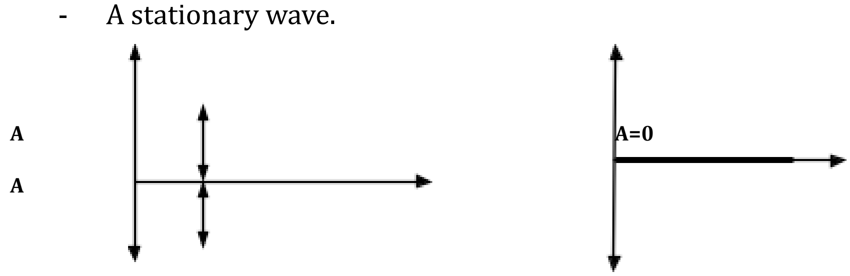
The waves are in phase and superimpose to produce a wave with a greateramplitude.

- A smaller wave is formed i.e. destructive interference.



The waves are out of phase with a phase difference of 180^0. Since they have different amplitudes, they superimpose to form a wave with a smaller amplitude.

- A stationary wave.



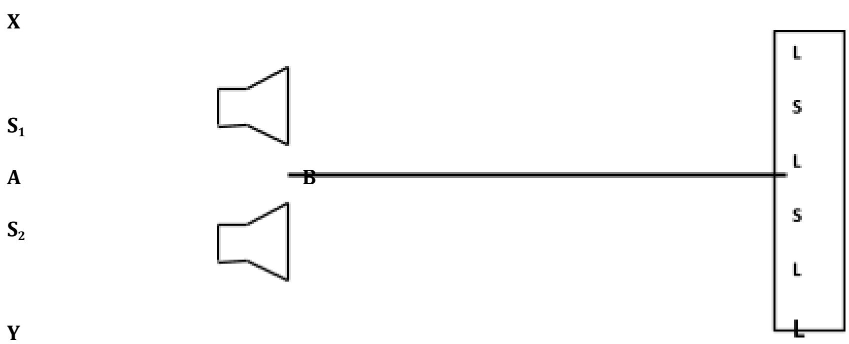
When the two waves which are out of phase with a phase difference of 180^0 superimpose, the result is a stationary wave having a zero amplitude.

Interference is a product of the **principle of superposition** which states: for two waves travelling in at a given point in the same medium, the resultant effect is the vector sum of their respective displacements.

- Interference of water waves can be shown by setting up two spherical dippers in a ripple tank which simultaneously generate waves. Alternating dark and bright radial lines will be observed on the screen representing regions of constructive and destructive interference respectively.

For interference to occur there ought to be a coherent source i.e. a source that generates waves of the same frequency and wavelength, equal or comparable amplitudes and having a constant phase difference.

- Interference of sound waves can be investigated by the set up below:

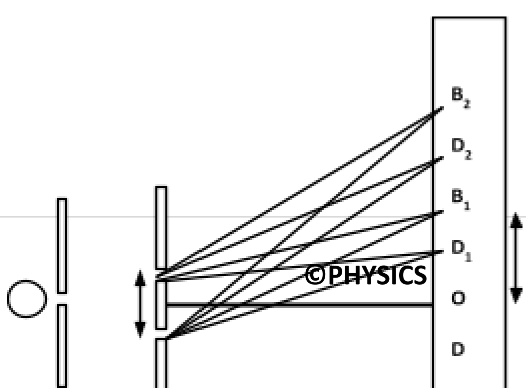


Two loudspeakers S_1 and S_2connected to an audio-frequency generator act as a coherent source. To an observer walking along a straight path XY, alternating loud and soft sound is heard. Along the line AB, a constant loud sound will be heard.

The regions with loud sound represent areas of constructive interference while the regions with soft sound represent areas of destructive interference. When the frequency of the signal is increased, the separation between the alternating loud and soft sound is reduced i.e. more close. Note that for a signal of any velocity, the higher the frequency the shorter the wavelength.

If instead the loudspeakers are connected such that the waves generated by one loudspeaker are exactly out of phase with those from the other, then all points along XY will have destructive interference and hence soft sound is heard throughout.

- Interference of light waves- this can be demonstrated by the Young's double slit experiment. Two narrow and very close slits S_1 and S_2 are placed infront of a monochromatic light source.



S_1 y

Light source S_2 x

Screen

d

The light waves from the two slits undergo diffraction and superimpose as they spread out. A series of alternating bright and dark fringes are observed on the screen. The bright fringes are due to constructive interference while the dark fringes are due to destructive interference. However, along the central line through the centre of the slits and point O, it is bright throughout.

At O, the path difference of the two waves is zero since $S_1O=S_2O$. Moving upwards or downwards to the first bright fringe, the path difference is equivalent to one wavelength;

i.e $\mathbf{S_2B_1-S_1B_1= 1\lambda}$

At D_1, the path difference is equivalent to half a wavelength;

$S_2D_1-S_1D_1= 1/2\lambda$

Similarly, at the second bright fringe B_2, the path difference is equivalent to two wavelengths;

i.e $\mathbf{S_2B_2-S_1B_2= 2\lambda}$

And $\mathbf{S_2D_2-S_1D_2= 3/2\lambda}$

Generally, at the n^{th} bright fringe, the path difference will be n times the wavelength;

$\mathbf{S_2B_n-S_1B_n= n\lambda}$

The wavelength of the light used can also be determined from the expression below:

$\mathbf{\lambda= xy/d,}$

Where x- the slit separation,

y- Distance between successive bright fringes and

d- Perpendicular distance of the slits from the screen.

16.6: Stationary waves verses progressive waves

A progressive wave is a wave that continuously moves away from the source. When two progressive waves equal in amplitude and travelling in opposite directions superpose on each other, the resultant wave is referred to as **a stationary or standing wave**. It is a common occurrence in stringed instruments. When the string is plucked/played, a transverse wave travels along the string and is reflected back on reaching the other end of the string.

A AAAA

N NNNNN

Reflected wave

The points marked N are always at rest (zero displacement) and are called nodes while those marked A are where the wave has maximum amplitude (maximum displacement). They are called antinodes.

When two loudspeakers connected to the same audio-frequency generator are such that they face each other, then the two sound waves superpose to produce a stationary wave.

For two progressive waves to produce a stationary wave, the following conditions must be satisfied:

- They must be travelling in opposite directions.
- Must have same speed, frequency and same or nearly the same amplitudes.

The following table gives the comparison between a stationary and a progressive wave:

Stationary waves	Progressive waves
Do not move through the medium hence does not transfer any energy from the source.	Move through the medium transferring energy from the source to a point away.
The distance between successive nodes or antinodes is equal to 1/2λ.	The distance between successive crests or troughs is equal to the wavelength of the wave.
The amplitudes of particles between successive nodes are different.	The amplitudes of any two particles which are in phase are the same.

TOPIC 17: SOUND(form two)

17.1: Introduction

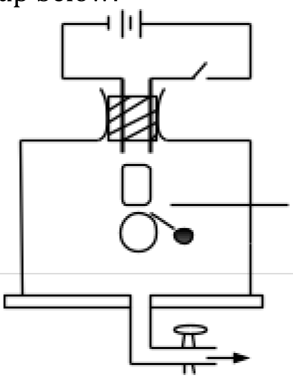
A sound may be defined as a form of energy produced by a vibrating object. Sound may be enjoyable if well organized such as in music. However, sometimes sound is a nuisance e.g noise pollution. Sound is a longitudinal wave. The human ear can be used to detect sound that is within the audible range i.e of frequency between 20Hz and 20kHz.

When an object vibrates, the air molecules around it are set into motion, collide forming compressions and rarefactions. Below are some common sources of sound:

- ✔ A vibrating thin wooden strip.
- ✔ A vibrating drum.
- ✔ A vibrating air column.
- ✔ A rotating cogged wheel and a card.
- ✔ Loudspeakers.
- ✔ A tuning fork.
- ✔ A guitar and other stringed instruments, etc.

17.2: Propagation of sound

Sound waves require a material medium to move through. They never travel in a vacuum. This can be illustrated by the set up below:



Electric bell

To vacuum pump

Initially when the bell is on the sound is heard from outside but as the air inside is sucked out, the sound gradually fades away. No sound is heard at all when all the air has been withdrawn. Hence, sound requires a material medium for its propagation.

Sound travels in all the three states of matter. It travels faster in solids, followed by liquids and slowest in gases. This is because the particles in solids are close together while they are far apart in gases.

17.3: Reflection of sound and speed of sound.

When a sharp sound falls on an obstacle, it is reflected back. Reflected sound is known as **an echo**. Sound waves obey the laws of reflection. When an incident sound wave and the reflected sound wave overlap, the original sound will appear prolonged. This effect is referred to as **reverberation**.

To minimize this effect, the minimum distance between the source of sound and the reflector should be 17m. It is for this reason that in tv and radio studios, the walls are made of absorbent materials to minimize the effect of echoes.

The idea of echo can be used to estimate the speed of sound in air. It entails measuring the distance between the source of the sound and the reflector and the average time interval between the production of the sound and the hearing of the echo;

Thus, speed of the sound= total distance covered/time interval.

Suppose the distance between the source of sound and the reflector is **x** and the time interval between the production of the sound and the hearing of the echo is **t**, the speed of the sound is expressed as;

Speed v=**2x/t.**

Example 17.1

1. A man makes a loud sound and hears an echo of the sound 1.2s later. If the speed of sound in air is 330m/s, calculate the distance between the man and the object reflecting the sound.

Let the distance between the man and the reflector be **x,** then;

Speed v= 2x/t= 2x/1.2 =330m/s

x=(1.2*330)/2 =198m

2. A student standing between two cliffs and 500m from the nearer cliff clapped his hands and heard the first echo after 3s and the second echo 2s later. Calculate:

a) the speed of the sound in air

Speed v=(2*500)m/3s =333.3m/s.

b) the distance between the cliffs

Let the distance between the student and the other cliff be **x.**

Then, 2x/(3+2)=333.3

x=(333.3*5)/2 =833.25m

Therefore, the distance between the two cliffs is= (500+833.25)= 1333.25m

17.4: Factors affecting speed of sound in air

The speed of sound in air under normal conditions is about 330m/s. However, this can change depending on the prevailing factors as illustrated below:

- **Temperature of the air:** warm air is lighter than cold air. An increase in the temperature of air causes an increase in the speed of the sound.

- **Wind:** if wind blows in the direction of the sound, its speed will be raised while if it blows against the sound, the speed will be reduced.

- **Humidity:** the speed of sound increases with humidity. Moist air having water vapour in it is less dense compared to dry air. An increase in the humidity of the air makes the air lighter and hence the speed of the air is increased.

- **Density of air:** the greater the density of air the lower its speed.

Not that as long as the temperature of air is constant, there is no effect of changing pressure on the speed of sound.

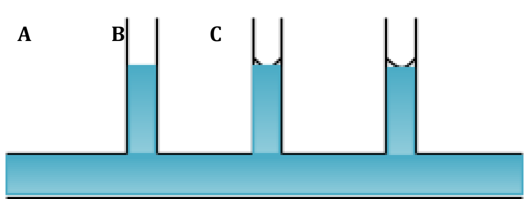
TOPIC 18: FLUID FLOW

18.1: Introduction

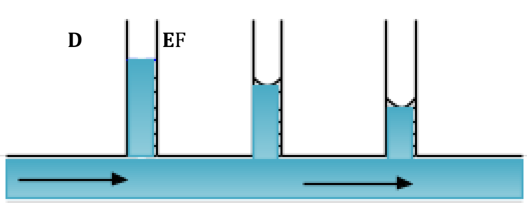
A fluid refers to liquids and gases. The study of flow of fluids is called **hydrodynamics**. When an object moves through a fluid, it experiences some resistance in terms of friction. The friction experienced in fluids is referred to as **viscosity** or **viscous drag**.

However, in this topic viscosity is assumed to be negligible. Certain shapes experience a lot of resistance when moving in fluids. However, we have some shapes which penetrate easily in fluids like air. Such shapes are called streamlined bodies. They include shapes of saloon cars, ships, rockets, fish, birds etc.

When a fluid is at rest in a horizontal tube, it exerts the same pressure at all points along the tube. This can be illustrated by connecting similar vertical tubes along the horizontal one.



The level of the liquid is the same in all the tubes. This indicates that the pressure acting on the liquid is the same all through. However, when the liquid is set to flow its level in the tubes will vary. This means that the pressure exerted by the liquid varies along the horizontal tube.



As the liquid flows through the tube, there exists frictional force between the molecules of the liquid and the wall of the tube. As a result, the pressure exerted by the liquid reduces as the liquid continue to flow. Hence the pressure of the liquid is greater at the beginning of flow and lower at the end of flow.

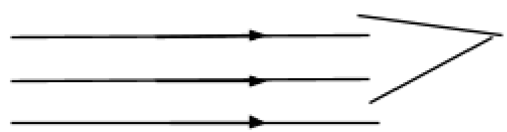
18.2: Types of flow

There are three main types of flow as far as fluid flow is concerned:

18.2.1: Streamline flow

In this type of flow, the fluid molecules in all the layers travel in the same direction with the same speed. The molecules move along paths called streamlines. In streamline flow, the streamlines are parallel to each other.

Streamlines



For a fluid to have streamline kind of flow, the following conditions are necessary:

-The fluid must be **non-viscous** so that the velocity of the fluid at all points along the tube is the same.
-The fluid must be **incompressible** so that the density of the fluid is the same at all points.

18.2.2: Turbulent flow

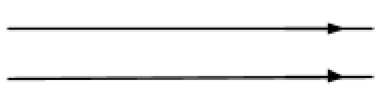
When an obstacle is placed on the path of the streamlines, the molecules will hit the barrier, lose some of its energy and may then change direction. This can lead to a disturbance of some sought resulting in formation of eddies or whirls. This is referred to as **turbulent flow**.

Obstacle

18.2.3: Laminar flow

This is where all the molecules in each layer move with the same velocity but different from that of the other layers i.e. each layer has its own velocity.

V_1



V_2

18.3: Rate of flow

This is the volume of fluid passing through a given region per unit time. It is also known as the **volume flux**.

Rate of flow= volume of fluid/time interval

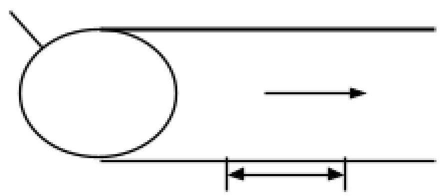
The SI Unit of rate of flow is the cubic metre per second (m^3/s). The fluid should not exceed a certain velocity called the critical velocity. If the critical velocity is exceeded, turbulence occurs.

Suppose a fluid takes a time t(s) to flow through a pipe of cross-section area A and length L, then:

The velocity Ѵ of the fluid=length L/time t.

Therefore, L=Ѵt …………………………………….. (i)

Area A



L

Also, the volume V of the fluid that passes through the region L=cross-section area A*length L.

Thus V=AL ……………………………………………… (ii)

Substituting equation (i) in (ii), we get:

V=AL=AѴt

Therefore, the rate of flow= volume flowing/timed interval

= V/t = AѴt/t

= AѴ

Hence the rate of flow is the product of the cross-section area and the velocity with which the fluids flows.

If the density of the fluid is ρ, then the **mass flux** is given by the product of the volume flux and the density of the fluid;

Mass flux= volume flux*density of the fluid= AѴρ.

Mass flux may be defined as the mass of the fluid passing through a given point per unit time.

Example 18.1

1. 600g of water flows through a tube of uniform cross-section area of $20cm^2$ in 10s. If the density of water is $1000kgm^{-3}$, calculate the rate of flow of the water.

Rate of flow= volume/time

But volume= mass/density =0.6kg/1000kgm^{-3}=6.0*10^{-4}m^{3}

Therefore, rate of flow=(6.0*10^{-4}m^{3})/10s

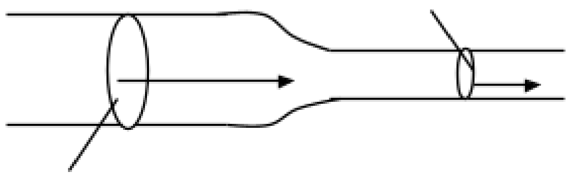
=6.0*10^{-5}m^{3}/s

18.4: The equation of continuity

Suppose a fluid of density ρ flows through a tube of varying cross-section as shown below:

Cross-section area A_2

V_1 V_2



Cross-section area A_1

Then, if the flow is streamline the rate of flow of the fluid will be the same at all points of the tube.

Thus, the rate of flow in the wider region= A_1V_1.

Similarly, the rate of flow in the narrower region= A_2V_2.

Therefore, for a streamline flow; $A_1V_1= A_2V_2$.

Generally, for a streamline flow when a fluid flows from a region of cross-section area A_1 with a velocity V_1 into a region of cross-section area A_2 with a velocity V_2, then:

$A_1V_1= A_2V_2$= a constant.

This equation is called the equation of continuity. The equation only holds under the following conditions:

- ✔ When the flow is streamline.
- ✔ When the fluid is non-viscous.
- ✔ When the fluid is incompressible.

Since the mass is directly proportional to volume, the mass flux will also be constant;

Therefore, $\mathbf{A_1V_1\rho=A_2V_2\rho}$.

Example 18.2

1. In the figure below, calculate the speed of water having a streamline flow at point P. The diameter of the pipe at P is 2cm and 6cm at Q. The speed of water at Q is 0.3ms^{-1}.

P

Q

$A_P V_P = A_Q V_Q$

$\pi \times 0.01^2 \times V_P = \pi \times 0.03^2 \times 0.3$

$V_P = (0.03^2 \times 0.3)/0.01^2 = 2.7 m/s$

2. A lawn sprinkler has 40holes each of cross-section area 2.0 x10^{-2}cm^2. It is connected to a horse pipe of cross-section area 1.6cm^2. if the speed of water in the horse pipe is 1.2m/s, calculate:

   a) The rate of flow in the horse pipe.

   Rate of flow= AV

   $=(1.6 \times 10^{-4}) \times 1.2 m/s = 1.92 \times 10^{-4} m^3/s$

   b) The speed at which water emerges from the holes.

   $A_1 V_1 = A_2 V_2$

   $1.92 \times 10^{-4} m^3/s = (2.0 \times 10^{-2} \times 10^{-4}) \times V_2$

   $V_2 = (1.92 \times 10^{-4} m^3/s)/(2.0 \times 10^{-2} \times 10^{-4} m^2) = 2.4 m/s.$

18.5: Bernoulli's principle

The principle states: provided a fluid is non-viscous, incompressible and its flow streamline, an increase in its velocity causes a corresponding decrease in the pressure it exerts.

Recall: The equation of continuity, $A_1 V_1 = A_2 V_2$.

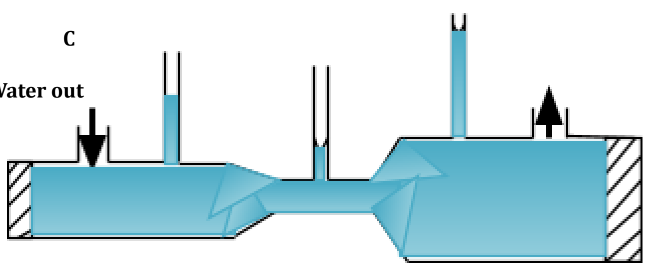
Rearranging the equation, we get;

$A_1/A_2 = V_2/V_1$

Hence velocity of the liquid is inversely proportional to the cross-section area i.e the smaller the cross-section area the larger the velocity and vice versa. This can be observed when water is allowed to flow through a tube of varying cross-section area as shown below:

A B C

Water in Water out



The velocity is high in the narrower region and low in the wider region. Thus the pressure of the liquid is greater in the wider region compared to the narrower region. This is shown by the height of the liquids in the vertical tubes i.e the greater the height the larger the pressure (**hint: P=hρg**).

Suppose a fluid of mass **m** and density **ρ** flows through a tube with a velocity **V** at a point where its pressure is **P**, then:

The kinetic energy per unit volume of the fluid=K.E/Volume V $=mV^2/2V$

But $m/V=\rho$

Therefore, K.E per unit volume=$\mathbf{\frac{1}{2}\rho V^2}$

Similarly, potential energy per unit volume of the fluid= P.E/Volume V

$$=mgh/V$$

Therefore, P.E per unit volume=$\mathbf{\rho gh}$

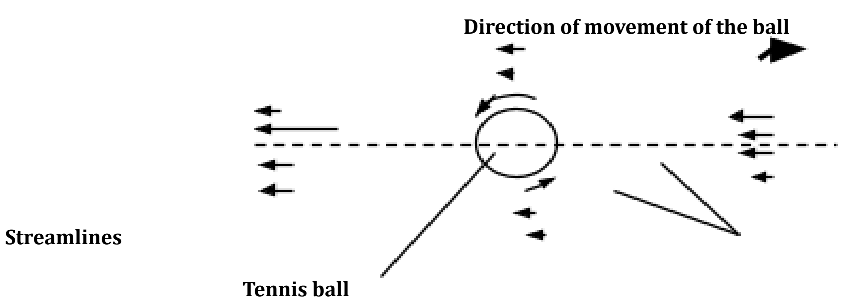
For a fluid which is incompressible, non-viscous and its flow streamline, the sum of the pressure P, kinetic energy per unit volume and potential energy per unit volume is a constant;

$\mathbf{P+\frac{1}{2}\rho V^2+ \rho gh=}$ **a constant**

This is the alternative statement of Bernoulli's principle.

The following are some of the effects of Bernoulli's principle:

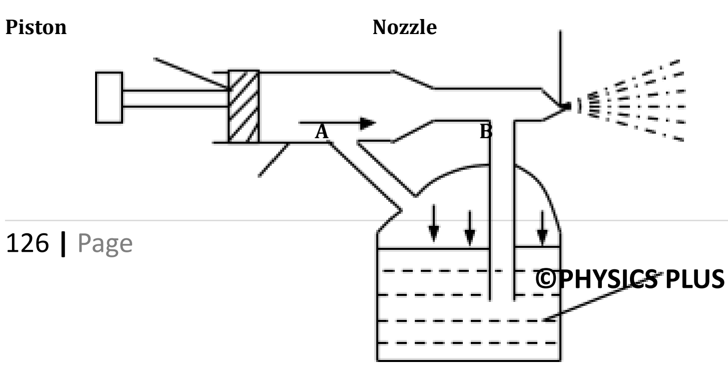
- When a light piece of paper is held in front of the mouth and some air blown above it, the paper is observed to be lifted upwards. This is because the velocity of air above the paper increases causing a corresponding decrease in pressure there. A pressure difference is thus created and the resultant upward force lifts the paper upwards.
- When a light tennis ball of negligible weight is made to spin in air, it is observed to curve away from its initial path.



As the ball spins, it drags air along with it. Suppose the ball spins in the direction shown, then the velocity of air on the lower side will be lower since it opposes the streamlines on that side. On the upper side, the velocity of air is higher. By Bernoulli's principle, there will be greater pressure acting from below the ball. The resultant upward force makes the ball to curve away from its original path.

18.6: Applications of Bernoulli's principle

18.6.1: A spray gun

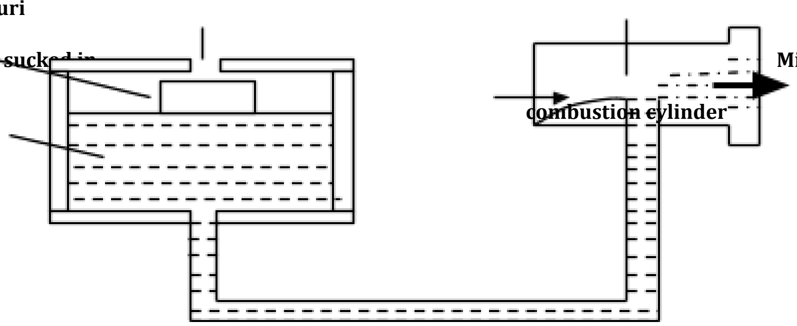


Barrel **P**

Liquid

When the piston is pushed forward, a portion of the air in the barrel is forced down into the reservoir through A while the remaining portion flows towards the nozzle. The air has higher pressure in the wider region and lower pressure in the narrower region. Thus the air forced through A exerts a lot of pressure on the surface of the liquid insecticide in the reservoir. The resultant upward force pushes the liquid up the tube B. this is then blown to the nozzle where it leaves with a high velocity.

18.6.2: Carburetor



With the help of the pressure exerted by the floater, petrol is forced into the venturi where it mixes with fast-moving air. The mixture is then drawn into the combustion cylinder of the engine where it is burnt to help operate the car.

18.6.3: Aerofoil

They are structured in such a way that the air flowing above them does so at very high speeds than that below them. Thus the resultant upward force acting on them helps them to rise to greater heights. Examples of aerofoil include aircraft wings and helicopter propellers.

18.7: Dangers of Bernoulli's effect

- ✔ Blowing of roof tops- when a fast moving air passes above a roof top, the air pressure above the roof will be higher than that below it. Thus the resultant upwards force acting on the roof may blow it off if it is large enough.
- ✔ Car accidents- when a small car and a large truck both travelling at high speeds and in opposite directions overtake each other, the small car is likely to be dragged onto the truck. The speed of air between the car and the truck is very high and hence a low pressure exists between them. The resultant inward force acting them draws them closer together. The car being lighter has a higher chance of being dragged onto the truck, causing an accident.

TOPIC 19: LINEAR MOTION(form three topic one)

19.1: Introduction

The study of motion is divided into two areas namely kinematics and dynamics. Kinematics deals with the motion aspect only while dynamics deals with the motion and the forces associated with it.

There are three common types of motion:

- Linear or translational motion.
- Circular or rotational motion.
- Oscillatory or vibrational motion.

In this topic, we concentrate on linear motion.

Note that all motion is relative i.e the state of a body; at rest or in motion, is **ONLY** true with respect to the observer's position.

19.2: Terms associated with linear motion

- **Distance-** is the length of the path covered by a body. It only gives the magnitude but no direction i.e it is a scalar quantity.
- **Displacement-** is the distance through which a body travels in a specified direction. It is a vector quantity.

Both distance and displacement are measured in metres.

- **Speed-** is the distance covered per unit time.

Speed= distance/time.

- **Velocity-** is the rate of change of displacement.

Velocity= displacement/time.

It is a vector quantity.

When the rate of change of displacement is non-uniform, we talk about average velocity;

Average velocity= total displacement/total time.

Both speed and velocity are expressed in metre per second (m/s).

- **Acceleration-** is the rate of change of velocity.

Thus, Acceleration= change in velocity/time interval = (final velocity v - initial veolicity u)/time.

Acceleration is measured in metre per square second (m/s^2).

If the velocity of a body decreases with time, its acceleration becomes negative. A negative acceleration is referred to as **deceleration or retardation**.

Example 19.1

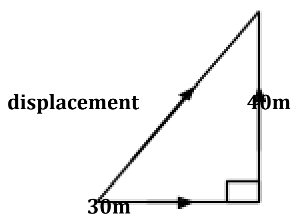
1. A body covers a distance of 10m in 4seconds, rests for 10seconds and finally covers a distance of 90m in 6seconds. Calculate its average speed.

Average speed= total distance/time= (10m+90m)/(4s+10s+6s)

= 100m/20s =5m/s.

2. A body moves 30m due east in 2seconds, then 40m due north in 4seconds. Determine its:

a) Average speed.



Average speed= total distance/time= (30m+40m)/(2s+4s)

=70m/6s =11.67m/s.

c) Average velocity.
d) Average velocity= total displacement/time =50m/6s

=8.33m/s.

3. A body is made to change its velocity from 20m/s to 36m/s in 0.1s. What is the acceleration produced?

a= (v-u)/t =(36m/s – 20m/s)/0.1s

=160m/s^2.

4. A particle moving with a velocity of 200m/s is brought to rest in 0.02s. What is the acceleration of the particle?

a= (v-u)/t =(0m/s-200m/s)/0.02

= -200/0.02 = -10,000m/s^2.

19.3: Motion graphs.

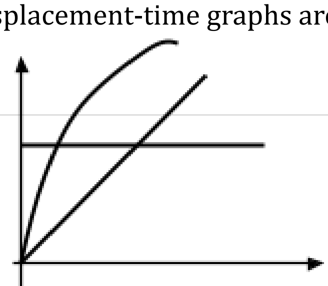
There are two categories; displacement-time graphs and velocity time graphs.

19.3.1: Displacement-time graphs

The slope of a displacement-time graph gives the velocity of the body.

The various displacement-time graphs are as illustrated below:

D **B**



Displacement **A**

(m) **C**

Time (s)

Graph A: the body is at rest i.e there is no change in displacement as time changes. The slope of the graph and hence the velocity is zero.

Graph B: the body moves with a uniform or constant velocity.

Graph C: the graph becomes steeper with time. The steeper the slope, the higher the velocity. Thus velocity of the body increases with time. The body is therefore accelerating.

Graph D: the graph becomes less and less steep with time i.e the body has a higher velocity at the beginning and decreases with time. Therefore, the body is said to be decelerating.

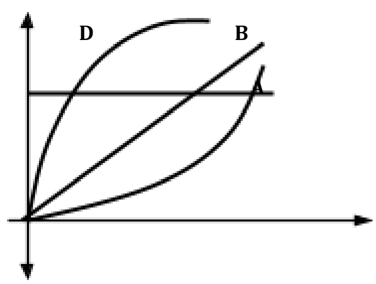
19.3.2: Velocity-time graphs

The slope of a velocity-time graph gives the acceleration of the body. Note that the area under a velocity-time graph gives the distance covered by the body.

The diagram below shows the possible velocity-time graphs:

Velocity(m/s)

C



Time(s)

Graph A: the velocity remains constant/uniform as time increases. The slope of the graph and hence the acceleration of the body is zero.

Graph B: the velocity changes uniformly with time. The body moves with a uniform/constant acceleration.

Graph C: the acceleration is lower where the graph is gentle and higher where the graph is steeper. Hence the acceleration of the body increases with time.

Graph D: in this case, the graph is steeper at the beginning and becomes gentle with time. Hence the acceleration of the body decreases with time.

19.4: Determination of velocity and acceleration

Two methods are applicable here:

Method 1: Using appropriate instruments e.g a tape measure and a stop watch to measure the displacement of a body and the duration then applying the formula;

Velocity= total displacement/time taken.

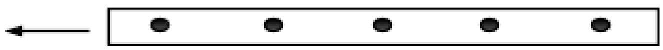
Method 2: Using a ticker-timer. It is used to measure velocity of a body specifically over short distances. It consists of an electronic vibrator which makes dots on a moving paper tape attached to the object whose velocity is being measured. The dots are made at a certain set frequency. For instance, a ticker-timer whose frequency is 50Hz makes dots at intervals of 0.02s. The time interval between successive dots is referred to as a **tick**.

The spacing between the dots depends on the manner in which the body is moving i.e moving at constant velocity or with increasing velocity or decreasing velocity. Generally, the dots are close together when the velocity is low and wide apart when the velocity is high. There are three possible patterns that can be obtained by a ticker-timer as illustrated below:

a) Moving at constant velocity.

The dots are equally or evenly spaced.

Direction of motion of the body



b) Moving with increasing velocity (accelerating).

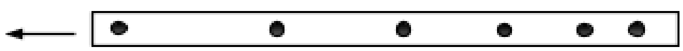
The spacing between the dots is initially small but increases away.

Direction of motion of the body

c) Moving with decreasing velocity (decelerating).

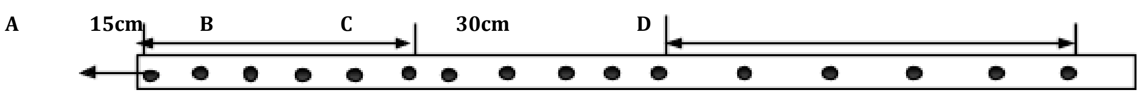
The spacing between the dots is initially large but decreases away.

Direction of motion of the body



Example 19.2

1. A paper tape was attached to a moving trolley and allowed to run through a ticker-timer. The figure below shows a section of the tape.



If the frequency of the ticker-timer is 100Hz, determine:

a) The velocity between AB and CD.

1tick= 1/100 =0.01s

V_{AB}= 15cm/(5ticks*0.01s) =15cm/0.05s

=300cm/s

V_{CD}=30cm/(5ticks*0.01s) =30cm/0.05s

=600cm/s

b) The acceleration of the trolley.

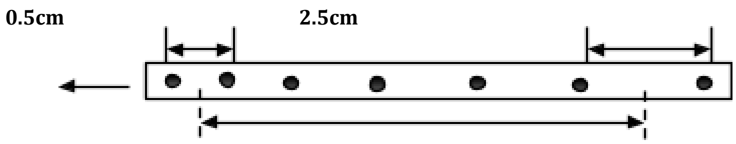
Note that the velocities calculated in (a) above are average velocities and as such are taken to be the velocities at the midpoints of AB and CD respectively. Hence, the time taken for the change in velocity is the time between the midpoints of AB and CD.

A 15cm B C 30cm D

$V_{AB}$$\Delta$t=10ticks*0.01=0.10s V_{CD}

Therefore, acceleration=(V_{CD}- V_{AB})/Δt= (600-300)cms^{-1}/0.10s =3000cms^{-2}.

2. The figure below represents part of a tape pulled through a ticker-timer by a trolley moving down an inclined plane. If the frequency of the ticker-timer is 50Hz, calculate the acceleration of the trolley.



Δt=5ticks*0.02s=0.10s

Note that 1tick=1/50 =0.02s.

Initial velocity u =0.5cm/0.02s= 25cms^{-1}

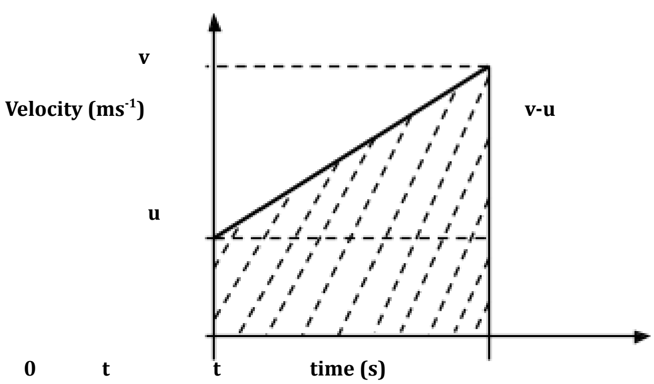
Final velocity v =2.5cm/0.02s= 125cms^{-1}

Hence, acceleration= (v-u)/Δt=(125-25)cms^{-1}/0.10s

=1000cms^{-2}

19.5: Equations of linear motion

There are three equations governing linear motion. Consider a body moving in a straight line from an initial velocity u to a final velocity v(u, v$\neq$0) within a time t as represented on the graph below:



The slope of the graph represents the acceleration of the body;

Acceleration, a=(v-u)/t.

Therefore, v=u+at…………………………………. i.

This is the first equation of linear motion.

The area under the graph (area of a trapezium) gives the displacement of the body.

Hence, displacement s= ½(sum of // sides)*perpendicular height between them.

$s= \frac{1}{2}(u+v)t$.

But $v=u+at$,

Therefore, $s=\frac{1}{2}\{u+(u+at)\}t$

$s=\frac{1}{2}(2u+at)t$

Hence, $s=ut+\frac{1}{2}at^2$………………………………. ii.

This is the second equation of linear motion.

Also, rearranging equation i, we have $t=(v-u)/a$. substituting this in equation ii, we obtain;

$s=ut+\frac{1}{2}at^2=u\{(v-u)/a\}+\frac{1}{2}a\{(v-u)/a\}^2$.

$s=u(v-u)/a + a(v-u)^2/2a^2= u(v-u)/a + (v-u)^2/2a$

$s= \{2u(v-u) + (v-u)^2\}/2a = \{2uv-2u^2+v^2+u^2-2uv\}/2a$

$s= \{v^2-u^2\}/2a$

$2as= v^2-u^2$

Hence, $v^2=u^2+2as$ ……………………………….. iii.

This is the third equation of linear motion.

The three equations hold for any body moving with uniform acceleration.

Note that for a body which is retarding, the acceleration a is given a negative sign.

Example 19.3

1. A particle travelling in a straight line at 2m/s is uniformly accelerated at $5m/s^2$ for 8seconds. Calculate the displacement of the particle.

$s=ut+\frac{1}{2}at^2=(2*8)+(\frac{1}{2}*5*8^2)$

$=176m$.

2. An object accelerates uniformly at $3ms^{-2}$. It attains a velocity of 19m/s in 5seconds.

a) What was its initial velocity?

$v=u+at$

$u= 19-(3*5)= 19-15 =4m/s$.

b) How far does it travel during this period?

$s=ut+\frac{1}{2}at^2= (4*5)+(\frac{1}{2}*3*5^2)= 57.5m$

3. A car travelling at 20m/s decelerates uniformly at $4m/s^2$. In what time will it come to rest?

v=u-at, (a is negative since the body is decelerating).

0=20-4t

t=20/4 =5seconds.

19.6: Motions under the influence of gravity

These include free fall, vertical projection and horizontal projection. The three equations of linear motion hold for motions under the influence of gravity.

19.6.1: Free fall

A body falling freely in a vacuum starts from an initial velocity zero and accelerates at approximately $9.8ms^{-2}$ towards the centre of the earth. This is called the acceleration due to gravity **g**. In this case, the air resistance is assumed to be negligible. Note that in a vacuum, a feather and a stone released from the same height will take the same amount of time to reach the surface of the earth.

Therefore, in the three equations of linear motion u=0m/s, s=h and a=g. thus the three equations become:

- ✔ v=gt, (from v=u+at)
- ✔ $h=½gt^2$, (from $s=ut+½gt^2$)
- ✔ $v^2=2gh$, (from $v^2=u^2+2as$)

From the above equations:

- $v= (2gh)^{½}$, where v is the velocity of the body just before it hits the ground.
- $h=½gt^2=v^2/2g$, where h is the height through which the body falls.
- $t=v/g=(2h/g)^{½}$, where t is the time of flight.

Example 19.4

1. A hammer falls from the top of a building 5m high.

a) How long does it take to reach the ground? Take $g=10ms^{-2}$.

$h=½gt^2$

$5=½*10t^2$

$t=1^{½}=1s$

b) With what velocity does it strike the ground?

$v= (2gh)^{½}= (2*10*5)^{½}=10m/s$.

19.6.2: Vertical projection

When a body is projected vertically upwards, it decelerates uniformly due to gravity until its velocity reduces to zero at maximum height. After attaining the maximum height, the body then falls back with an increasing velocity.

The body must be given an initial velocity and attains a final velocity of zero at its maximum height. Note that the sign of ‘g’ is negative for a vertical projection. This is because the body moves against gravity.

Hence the three equations of linear motion become:

- ✔ $v=u-gt$, (from $v=u+at$)
- ✔ $h=ut-½gt^2$, (from $s=ut+½gt^2$)
- ✔ $v^2=u^2-2gh$, (from $v^2=u^2-2as$)

But at maximum height h_{max}, $v=0$. Thus, the three equations reduce to:

i. $gt=u$,

ii. $h=ut-½gt^2$

iii. $u^2=2gh$.

From equation (i), the time taken to attain the maximum height is given by;

$t=u/g$.

Similarly, the initial velocity u and the maximum height attained by the body h_{max} can be expressed as:

$u=gt=(2gh_{max})^{½}$

And $h_{max}=ut-½gt^2=u^2/2g$.

When the body finally falls back to its point of projection, the displacement of the body will be zero. Substituting this in equation (ii), we obtain;

$0=ut-½gt^2$

Therefore, $0=t(2u-gt)$

And $t=0$ or $t=2u/g$, where $t=0$ is the time at the start of the projection and,

t is this is the total time of flight i.e for both upward projection and fall back. Note that the total time of flight is twice the time taken to attain maximum height.

also, the velocity of the body just before hitting its point of projection as it falls back is the same in magnitude but in opposite direction to its initial velocity; $v=-u$.

Example 19.5

1. A bullet is shot vertically upwards and rises to a maximum height of 1000m. Calculate:

a) the initial velocity of the bullet,

$u=(2gh_{max})^{½}= (2*10*1000)^{½}=$

b) the total time of flight.

$t=2u/g=2*$

2. An object is released to fall vertically from a height of 100m. At the same time, another object is projected vertically upwards with a velocity of 40m/s.

a) Calculate the time taken before the two objects meet.

Let the time taken to meet be t. then, after a time t the distance covered by the object moving downwards will be; $s_d=½gt^2$, (since u=0).

$=½*10t^2=5t^2$

The distance covered by the object projected upwards after a time t will be;

$s_u=ut-½gt^2=40t-5t^2$

But $s_d+s_u=100m$

Therefore, $5t^2+40t-5t^2=100$

t=100/40=2.5s

b) At what height above the point of projection do they meet?

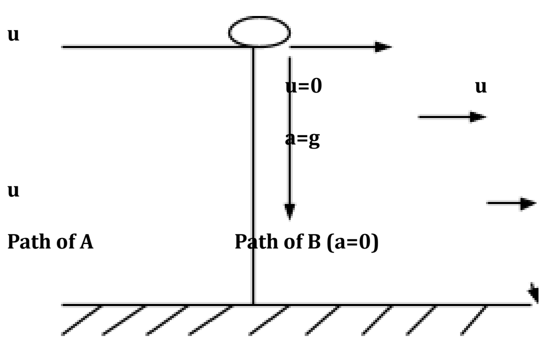
$s_u=ut-½gt^2=(40*2.5)-(½*10*2.5^2)$

=68.75m

19.6.3: Horizontal projection

If two objects A and B at a point some height above the ground are such that A is allowed to fall freely (vertically downwards) while is B given a horizontal projection with an initial velocity u, then both objects take the same duration to reach the ground. This is because both are acted on by the same gravitational force. The object on the horizontal projection moves with a constant velocity u. hence, the horizontal acceleration of the object is zero. For the object falling freely, the acceleration is equivalent to 'g' and the initial velocity u is zero. However, the object under horizontal projection will strike the ground some distance away from the point the other object strikes the ground. This horizontal distance covered by the object is referred to as the '**range R**'.

Note that both A and B will strike the ground with the same velocity.



Since a=0 for the horizontal projection, s=R=ut.

Also, the time taken to reach the ground in both cases is expressed as;

t=u/g.

Example 19.6

1. A stone is thrown at a velocity of 30m/s to the horizontal by a girl at the top of a tree whose height is 30m. Calculate:

a) the time taken for the stone to strike the ground.

Since both free fall and horizontal projection take the same duration;

$h=\frac{1}{2}gt^2$

$30=\frac{1}{2}*10*t^2$

$t=6^{\frac{1}{2}}=$

b) the velocity at which the stone strikes the ground.

u=0 (for free fall).

Therefore, $v=(2gh)^{\frac{1}{2}}=(2*10*30)^{\frac{1}{2}}$

$=$

2. A jet fighter on practice moving at a velocity of 100m/s released a bomb above the ground which hits the ground after 3s. Calculate:

a) the distance from the ground to the jet,

$h=\frac{1}{2}gt^2=\frac{1}{2}*10*3^2$

$=45m$

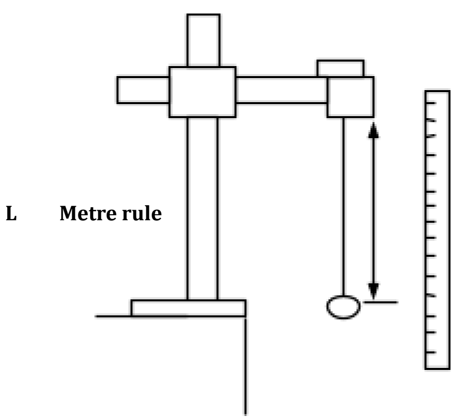
b) the horizontal distance from the target when the bomb is released.

R=ut=100*3

=300m.

19.7: Experimental determination of acceleration due to gravity.

This can be done as follows:



- Set the apparatus as shown in the diagram above. Set the length of the string at 30cm. note that the length l is measured from the centre of the bob.

- Displace the bob sideways through a small angle of about 10^0 and release it so as to oscillate.

- With the help of a stop watch, measure and record the time for ten oscillations (allow some little oscillations after release before timing). Repeat this step twice or thrice and determine the average time.

Hence calculate the period T(time for one oscillation).

- Repeat the above steps for l=40cm, 50cm, 60cm, 70cm and 80cm. complete the table below:

Length,l (cm)	**Time for 10oscillations, t(s)**				**Period, T(s)**	**$T^2(s^2)$**
	t_1	**t_2**	**t_3**	**$t=(t_1+t_2+t_3)/3$**		

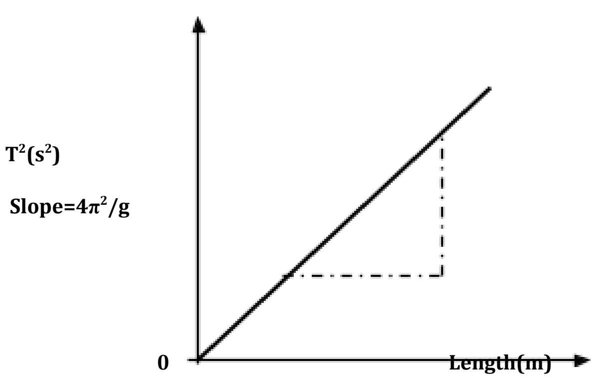
- plot a graph of T^2 against length l in metres.

Observations and conclusion

The frequency of oscillation increases with decrease in length of the string. A graph of T^2 against length l is a straight line through the origin.

Generally, a graph of T^2 against length for a simple pendulum satisfies the equation $T^2=4\pi^2 l/g$.

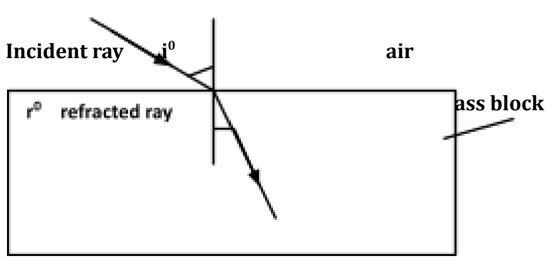
Hence, the slope of the graph above is equals to $4\pi^2/g$.



TOPIC 20: REFRACTION OF LIGHT

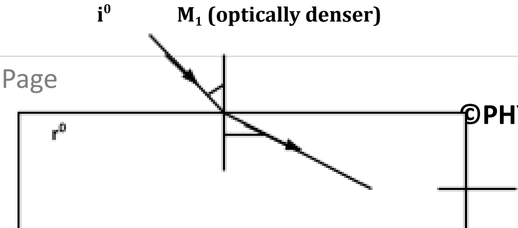
20.1: Introduction

Refraction refers to the bending of light when it passes from one medium into another of different optical density. This is because as light passes through different media its velocity changes. The bending occurs at the boundary or interface of the two media.



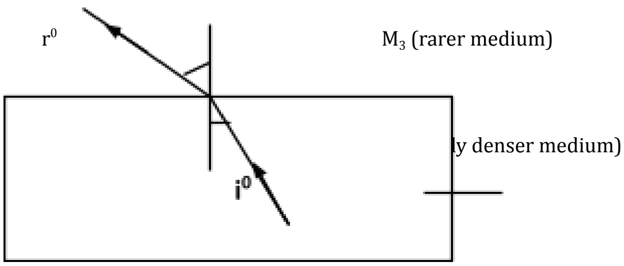
The refracted ray may bend away or towards the normal depending on the optical density of the second medium with respect to the first medium. Generally, a ray passing from an optically denser medium into a less optically dense (rarer) medium is bent away from the normal after refraction. If the ray passes from a rarer medium into an optically denser medium then it is bent towards the normal. It is easier to tell which medium is optically denser by simply comparing the angle between the incident ray and the normal and that between the refracted ray and the normal. The medium with a smaller angle (of incidence or refraction) is the optically denser medium.

(a) i^0 M_1 (optically denser)

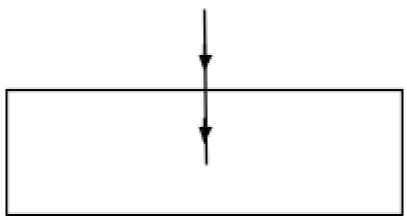


M_2 (rarer medium)

(b)



However, when the ray strikes the interface perpendicularly (normally) it passes undeviated (without bending). This is because the angle of incidence is zero.



In figure (b) above, only the direction of the light has been reversed leaving the angles the same. However, i now become r while r becomes i. The principle that makes it possible to reverse the direction of light keeping the sizes of the angles the rays make with the normal the same is called the principle of reversibility of light.

The study of refraction of light helps us understand the following common phenomena:

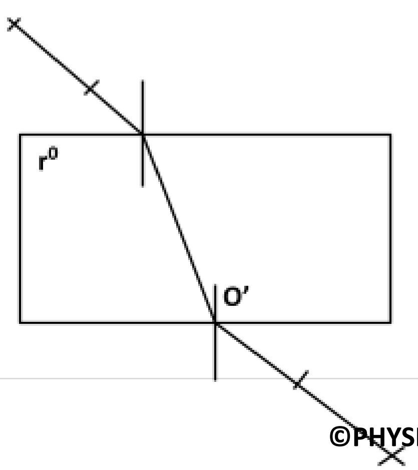
- Why a stick appears bent when part of it is in water.
- Why a coin at the base of a beaker of water appears nearer the surface than it actually is.
- Why the stars twinkle.
- Why the sun can still be seen sometimes before it rises or even after setting.
- Why the summer sky appears blue.
- The formation of the rainbow.

20.2: Refraction in glass

This can be investigated by the following steps:

P_1

$P_2 i^0$ O



e^0P_3

P_4

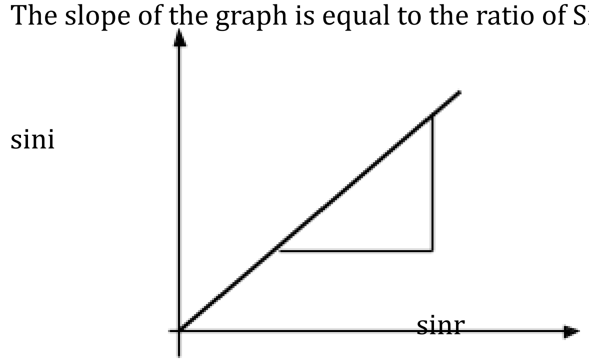
- Fix a white plain paper on a soft board using drawing pins. Place the glass block with its larger surface on the plain paper and trace its outline.
- Remove the glass block and then draw a normal through point O. Draw a line making an angle say $i=30^0$ with the normal as shown above.
- Replace the glass block onto the outline and stick two pins P_1 and P_2 along the line such that they are upright and about 6cm apart.
- From the opposite side of the block, view the two pins and stick two pins P_3 and P_4 such that the four pins appear on a straight line. Join the positions of P_3 and P_4 using a straight line and produce the line to meet the outline at O'.
- Draw another normal at O' and then join O to O'. Measure angles **r** and **e**.
- Repeat the above steps for other values of $i=40^0$, 50^0 and 60^0. Complete the table below:

Angle of incidence, i^0	30^0	40^0	50^0	60^0
Angle of incidence, r^0				
e^0				
Sin i				
Sin r				
Sini/Sinr				

- Plot a graph of Sin i against Sin r. determine the slope of your graph.

Observations

- The ratio of Sin i to Sin r is a constant.
- The graph of Sin i against Sin r is a straight line through the origin.
- The slope of the graph is equal to the ratio of Sin i to Sin r in the table.



20.3: The laws of refraction and refractive index

There are two laws of refraction:

1. The incident ray, refracted ray and the normal at the point of incidence all lie in the same plane.

2. **Snell's law**: it states that the ratio of sine of angle of incidence to the sine of angle of refraction is a constant for a given pair of media.

i.e. Sin i/Sin r = a constant.

The constant is referred to as the **refractive index, η** of the second medium with respect to the first medium. The first medium is that medium in which the incident ray is found while the second medium is that medium where the refracted ray is found. It is denoted as ${}_1\eta_2$.

Hence in 20.2 above, the ratio Sin i/Sin r is the refractive index of glass with respect to the air since the light passed from air into glass block.

However, when light passes from vacuum into another medium, it is referred to as absolute refractive index. Therefore for absolute refractive index, the angle of incidence **i**is found in a vacuum.

i.e. absolute refractive index= sin i(in vacuum)/sin r(in the second medium).

Recall:

${}_1\eta_2$=sin i/sin r

By the principle of reversibility of light, r now becomes i and i becomes r i.e. the incident ray is now found in the second medium.

Hence ${}_2\eta_1$=sin r/sin i

But sin r/sin i=1/(sin i/sin r)=1/${}_1\eta_2$

Therefore ${}_2\eta_1$=1/${}_1\eta_2$.

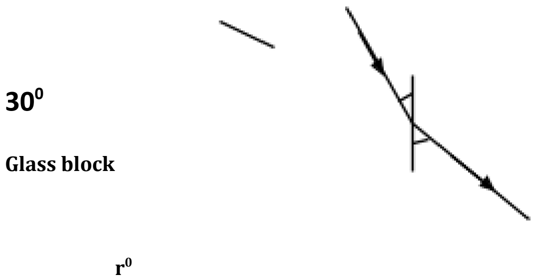
The table below shows some materials and their refractive indices:

Material	Refractive index
Ice	1.31
Crown glass	1.50
Water	1.33
Alcohol	1.36
Kerosene	1.44
Diamond	2.42

Note that the refractive indices given in the above table are with respect to air i.e. when light travels from air into the various media.

Example 20.1

1. In the figure below, calculate the angle of refraction r given that the refractive index of the glass is 1.50.



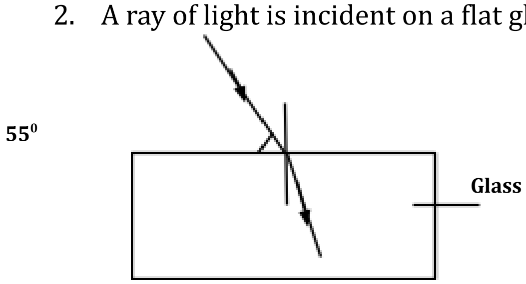
By the principle of reversibility of light;

sin r/sin 30^0= 1.50

sin r=1.50*sin 30^0

r= sin-1(1.50*sin 30^0)= 48.6^0.

2. A ray of light is incident on a flat glass surface as shown below:



Given that the refractive index of glass is 1.50, determine the angle of refraction for the ray of light.

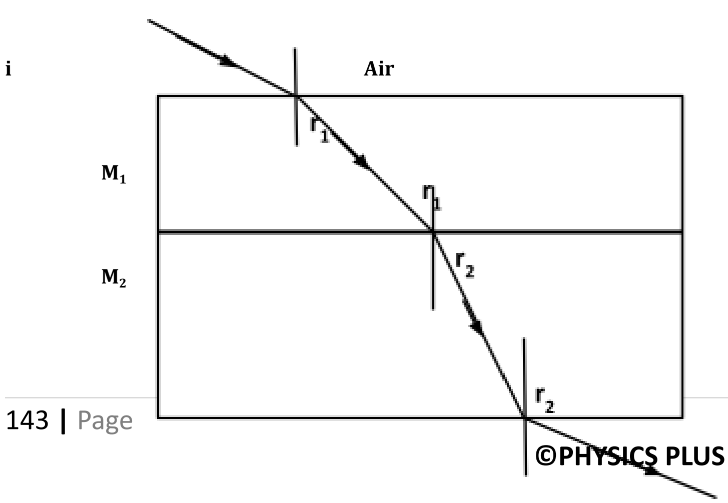
1.50= sin 35^0/sin r

Sin r= sin 35^0/1.50

r=sin-1(sin 35^0/1.50)= 22.48^0

20.3.1: Refraction through successive media

Consider a ray of light passing through a series of media as shown below:



i **Air**

Suppose the boundaries are parallel, then:

${}_a\eta_1$=sin i/sin r1………………………………. (i)

${}_1\eta_2$=sin r1/sin r2 …………………………… (ii)

${}_2\eta_a$=sin r2/sin i ……………………………… (iii)

By the principle of reversibility of light;

${}_a\eta_2$=sin i/sin r2……………………………. (iv)

Also, multiplying equations (i) and (ii), we get:

${}_a\eta_1 * {}_1\eta_2$= sin i/sin r_1 * sin r_1/sin r_2 =sin i/sin r_2.

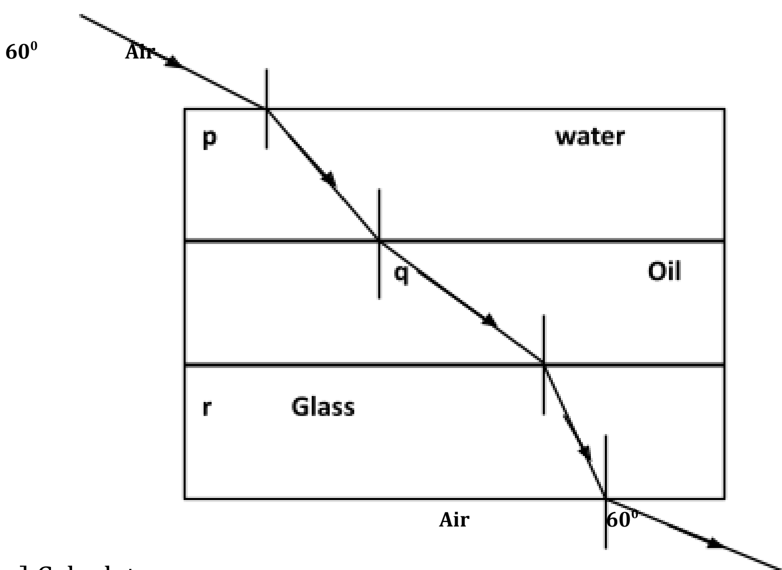
Thus ${}_a\eta_2 = {}_a\eta_1 * {}_1\eta_2$.

Generally, ${}_1\eta_k = {}_1\eta_2 * {}_2\eta_3 * \ldots\ldots\ldots\ldots * {}_{k-1}\eta_k$.

Example 20.2

1. A ray of light from air passes successively through parallel layers of water, oil, glass and then into air again. If the refractive indices of water, oil and glass are 4/3, 6/5 and 3/2 respectively and the angle of incidence in air is 60^0.

a) Draw a diagram to show how the ray passes through the multiple layers.



a] Calculate:

i) The angle of refraction in water.

4/3= sin 60^0/sin r

r=sin-1(3sin60^0/4)= 40.5^0

ii) The angle of incidence at the oil-glass interface.

$_{o}\eta_{g}=\sin q/\sin r$

By the principle of reversibility of light, $_{a}\eta_{g}=\sin 60^{0}/\sin r = 3/2$

r= $\sin^{-1}(2\sin 60^{0}/3) =35.27^{0}$.

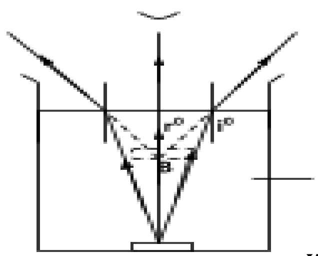
Also, $_{o}\eta_{g}= {_{o}\eta_{a}} * {_{a}\eta_{g}}=5/4$

Therefore, $5/4= \sin q/\sin 35.27^{0}$

$q=\sin^{-1}(5\sin35.270/4)=46.19^{0}$

20.4: Refractive index in terms of real and apparent depth

This is on the basis that when an object at the base of a container filled with water is viewed perpendicularly it appears closer to the surface than it actually is. Consider the figure below:



Water

A

From the figure, $_{w}\eta_{a}=\sin i/\sin r$.

Therefore, $_{a}\eta_{w}=\sin r/\sin i$.

Since the angles i and r are very small, sin i=tan i and sin r=tan r.

Therefore, by the principle of reversibility of light, $_{a}\eta_{w}=\sin r/\sin i =\tan r/\tan i = (CD/BC)/(CD/AC)$

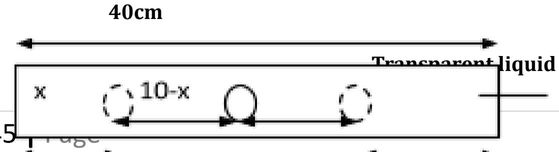
Thus $_{a}\eta_{w}$= AC/BC, where AC- real depth and BC- apparent depth.

Hence, refractive index of water= Real depth/Apparent depth.

When a graph of real depth against apparent depth is plotted, the graph obtained is a straight line through the origin and whose gradient is equal to the refractive index of the medium involved.

Example 20.3

1. In a transparent liquid container, an air bubble appears to be 12cm when viewed from one side and 18cm when viewed from the other side. If the length of the tank is 40cm, where exactly is the air bubble?



12cm **18cm**

Refractive index of glass= (12+x)/12 = (18+10-x)/18

x= 20/5 = 4cm.

Therefore, the bubble is 16cm in the liquid from the left-hand side.

2. A microscope is focused on a mark on a horizontal surface. A rectangular glass block 30mm thick is placed on the mark. The microscope is then adjusted 10mm upwards to bring the mark back to focus. Determine the refractive index of the glass.

${}_a\eta_g$=real depth/apparent depth= 30mm/20mm

=1.50

20.5: Refractive index in terms of velocity of light

Refraction occurs as a result of the different light velocity in different media. Basically, refractive index of any medium is the ratio of the velocity of light in a vacuum or air to the velocity of light in that medium;

η_m= velocity of light in vacuum/velocity of light in the medium .

Note that the velocity of light in a vacuum is $3.0*10^8$m/s.

Generally, ${}_1\eta_2$=velocity of light in medium 1/velocity of light in medium 2.

Example 20.4

1. The velocity of light in glass is $2.0*10^8$m/s. Calculate:

a) The refractive index of glass.

η_g= velocity of light in vacuum/velocity of light in glass= $(3.0*10^8)/(2.0*10^8)$ =1.50

b) The angle of refraction in glass for a ray of light incident at the air-glass interface at an angle of incidence of 40^0.

Sin 40^0/sin r =1.50

r=sin-1(sin40/1.50)= 25.4^0.

2. Calculate the speed of light in diamond of refractive index 2.4.

η_d=velocity of light in vacuum/velocity of light in diamond

2.4= $(3.0*10^8)/V_d$

$V_d=(3.0*10^8)/2.4 =1.25*10^8$m/s.

3. The speed of light in medium 1 is $2.0*10^8$m/s and in medium 2 is $1.5*10^8$m/s. Calculate the refractive index of medium 2 with respect to medium 1.

${}_1\eta_2=V_1/V_2$ = $(2.0*10^8$m/s)/ $(1.5*10^8$m/s)

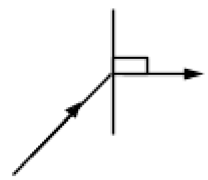
=1.33

20.6: Total internal reflection, critical angle and refractive index

As the angle of incidence in the denser medium increases the angle of refraction also increases. If this continues until the angle of refraction reaches 90^0, the angle of incidence is called the critical angle C. A critical angle is defined as the angle of incidence in the denser medium for which the angle of refraction is 90^0 in the less dense medium.

C

Air



By the principle of the reversibility of light,

${}_a\eta_g$= $\sin 90^0/\sin C = 1/\sin C$.

If the angle of incidence exceeds the critical angle, the light undergoes total internal reflection. This reflection obeys all the laws of reflection.

For total internal reflection to occur, two conditions must be satisfied, namely:

- ✔ Light must pass from an optically denser medium to a less optically dense medium.
- ✔ The angle of incidence in the denser medium must be greater than the critical angle.

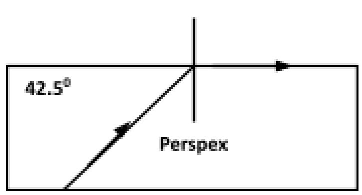
Example 20.5

1. Calculate the critical angle for glass whose refractive index is 1.50.

1.50= 1/sin C.

C = sin-1(1/1.50) =

2. The figure below shows the path of a ray light passing through a rectangular block of Perspex placed in air.



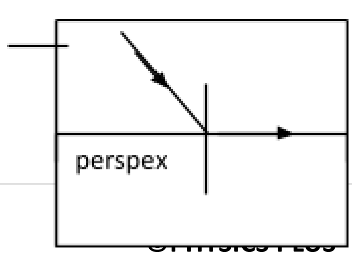
a) Calculate the refractive index of Perspex.

${}_a\eta_p = 1/\sin 42.5^0 = 1.48$

b) A ray of light now travels from a transparent medium of refractive index 2.4 into the Perspex as shown below:

Transparent material

C



Calculate the critical angle C.

${}_p\eta_m = \sin C/\sin 90^0 = {}_p\eta_a * {}_a\eta_m = (1/{}_a\eta_p) * {}_a\eta_m$

$= 1/2.4 * 1.48 = 1.48/2.4$

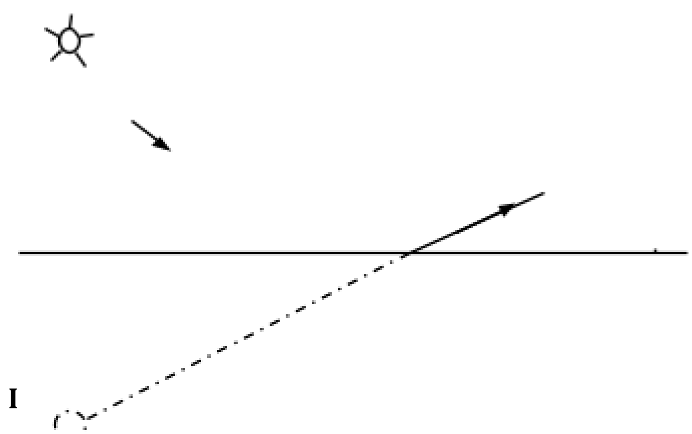
$C = \sin^{-1}(1.48 \sin 90^0/2.4) = 38.07^0$.

20.6.1: Effects of total internal reflection

- **Mirage**

On a hot day, the air above the ground is at a higher temperature than the layers above it. Thus the density of air increases with height above the ground. Denser air is optically denser than lighter one. Hence, a ray of light from the sun undergoes continuous refraction at the boundaries between any two layers of air with different temperatures. In each case, the ray bends away from the normal until the critical angle is achieved. Thereafter, the ray undergoes total internal reflection. An inverted image in the form of a pool of water is observed. This phenomenon is referred to as **mirage**.

Generally, mirage occurs as a result of continuous and progressive refraction at the air boundaries and total internal reflection. Mirage also occurs in cold regions but this time the ray of light curves upwards.



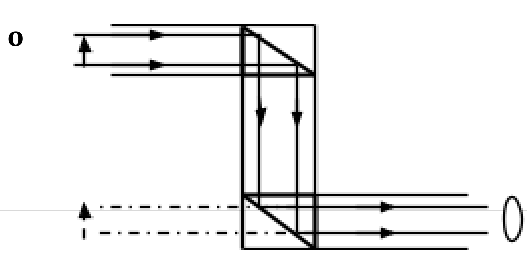
- **Atmospheric refraction**

The sun is sometimes seen before it actually rises or after it has set. This is because the light from the sun is refracted by the atmosphere towards the earth. (**Recall: the earth is spherical**).

20.6.2: Applications of total internal reflection

a) **A prism periscope**

It makes use of two right-angled isosceles prisms. The light from the object is inverted through 90^0 by the first prism and a further 90^0 by the second prism.



I

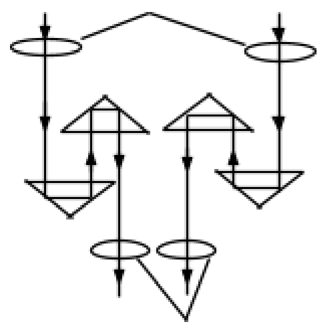
This periscope produces brighter images compared to those of the simple periscope in which a plane is used. The image formed is erect and virtual. A prism periscope has the following advantages over the simple periscope:

- ✔ Forms brighter and clearer images. A simple periscope produces many faint images besides the main image especially if the mirror is thick.
- ✔ Does not absorb the energy of the light. Plane mirrors absorb some light incident on them.
- ✔ Has a tough structure and thus does not easily wear. The painting on the plane mirror can wear out with time.

b) A prism binoculars

This device is used to reduce the distance between the eyepiece and the objective thereby reducing the length of the telescope. It forms an erect image.

Objective lenses

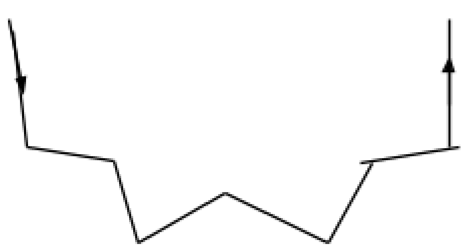


Eyepiece lenses

c) Optical fibre

It is a thin flexible glass rod made up of two parts; the inner part made of glass of higher refractive index and the outer glass coating of lower refractive index. When a ray of light enters the fibre at an angle greater than the critical angle, it undergoes a series of total internal reflection before it finally emerges from the other end. None of the light energy is lost in the process.

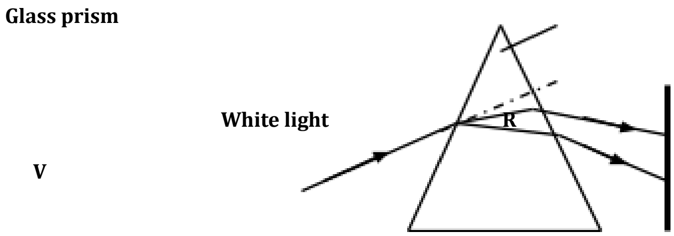
Optical fibres are used in medicine for viewing internal body organs (the endoscope) as well as in telecommunication. They are preferred to ordinary cables because they are light and thin and do not cause scattering of the signals.



20.7: Dispersion of light

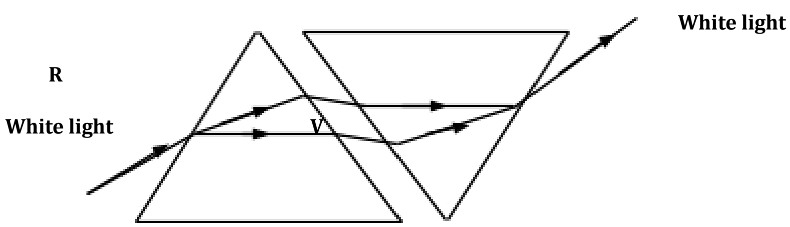
White light from the sun is made up of seven colours. They all travel with the same velocity in vacuum but their velocities vary in other transparent media like glass and water. Hence when a ray of white light travels from a vacuum into a glass prism, it is separated into its component colours ranging from red, orange, yellow, green, blue, indigo to violet. The spreading out of light into its constituent colours by another medium is called **dispersion**.

Pure light is called **monochromatic** light while an impure light like white light is referred to as **non-monochromatic or composite** light. Dispersion of light is illustrated by the diagram below:



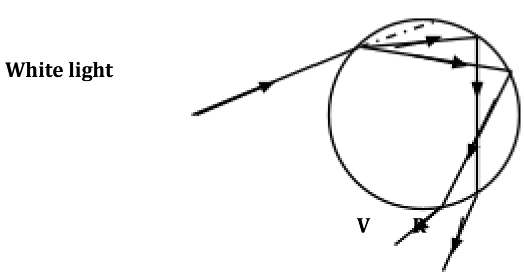
Red is least deviated while violet is the most deviated ray. Hence red light has the greatest velocity and violet the least velocity in glass. The coloured band produced is called a **visible spectrum**. The spectrum produced above is impure. In order to obtain a pure spectrum where each colour is distinct, an achromatic lens is placed between the screen and the prism.

When the seven sevencolours are recombined, a white light is obtained. This can be achieved by using a similar but an inverted prism.



20.8: The rainbow

When a ray of light passes through a water drop, a rainbow is produced. The water disperses the light into its constituent colours. Each colour then undergoes total internal reflection within the drop before it eventually emerges into air again.



TOPIC 21: NEWTON'S LAWS OF MOTION

21.1: Introduction

The laws governing the motion of a body are grouped into three. They are based on the effects of force on a body. Some of the effects of force on a body include:

- Force can make a stationary body to start moving.
- Can make a moving to stop.
- Can deform a body i.e. change its shape.
- Can change the direction of a moving body.
- Can change the speed of a moving body.

21.2: Newton's first law of motion

The law states: a body remains in its state of rest or uniform motion in a straight line unless acted upon by an external force. This explains the following common observations:

- Passengers in a bus are pushed forward when brakes are applied suddenly or backwards when a bus at rest takes off suddenly. Hence the fitting of seatbelts in vehicles.

- A coin placed on a cardboard on top of a glass tumbler drops into the tumbler when the cardboard is pulled sideways.

- Athletes run past the finish line of a race before they finally stop.

These observations show that bodies have an in-built reluctance to changes in their state of motion or rest. The tendency of a body to resist change in its state of rest or motion is called **inertia**. Hence Newton's first law of motion is also referred to as the **law of inertia**.

21.3: Newton's second law

This law states: the rate of change of momentum of a body is directly proportional to the resultant external force acting on the body and takes place in the direction of the force.

Moment of a body is defined as the product of its mass and velocity. Since velocity is a vector quantity, momentum is also a vector quantity having both magnitude (size) and direction.

Momentum P=mass m*velocity v

Hence the unit of momentum is the kilogram-metre per second (kgm/s). The direction of momentum is the same as that of the velocity. The change of momentum is therefore caused by a change in velocity.

Suppose the velocity of a body of mass **m** changes from an initial value **u** to a value **v** after a time **t**, then:

The initial momentum P_i=mu

The final momentum P_f=mv

The change in momentum= final momentum- initial momentum

Thus $\Delta P = P_f - P_i = mv - mu = m(v-u)$

Therefore, the rate of change of momentum= $\Delta P/t = m(v-u)/t$.

From the equations of linear motion, $(v-u)/t$ =acceleration **a**

Hence $\Delta P/t = ma$.

From the second law of motion, $F \alpha ma$.

And so the force F= mass m*acceleration a (F=ma).

Therefore, F=ma=m(v-u)/t

And Ft=m(v-u).

The product of the force and time is called **impulse**. It is a vector quantity since force is a vector quantity. The unit of impulse is the newton-second(Ns). Impulse is also equal to the change in momentum(mv-mu). Hence impulse can also be expressed in kgm/s.

Example 21.1

1. Two stones of mass 8kg and 4kg move with velocities 3m/s and 6m/s respectively. Compare their momentum.

P_{8kg}=mv =8*3=24kgm/s

P_{4kg}=mv =4*6=24kgm/s

Hence they have the same momentum.

2. A ball of mass 35g travelling horizontally at 20m/s strikes a barrier normally and rebounds with a speed of 16m/s. Find the impulse exerted on the ball.

   Impulse=Ft=m(v-u)= (0.035*20) – (0.035*-16)

   =1.26Ns

   Note that the two speeds are in opposite directions.

3. A kick that lasts 0.03s sends a ball of mass 0.65kg with a velocity of 15m/s northwards. Find:

a) The change in momentum of the ball.

   Note that the ball is initially at rest, i.e. u=0m/s.

   ΔP=mv-mu=(0.65*15)-(0.65*0)=9.75kgm/s

b) The average force exerted on the ball.

   F=m(v-u)/t =(9.75kgm/s)/0.03s)=325N

c) The displacement of the ball in 2 seconds.

   The upward acceleration of the ball is negative 10m/s^2.

   S=ut+1/2at^2=(15*0.03)+(1/2*-10*0.03^2)=10m/s.

21.4: Newton's third law

The law states: for every action there is an equal and opposite reaction. We look at the working of a lift in relation to the third law of motion in three situations:

a) **When the lift is at rest.**

This implies that the resultant force on the lift is zero i.e. action and reaction are equal in size. The force acting on the lift is the weight of the person standing in the lift. This is balanced by the reaction by the floor of the lift.

Therefore, weight mg=- reaction R,

Or simply; mg+R=0.

b) When the lift descends with an acceleration a

For the lift to move downwards, the weight of the occupant must be greater than the reaction by the floor of the lift. Therefore, the resultant force pulling the lift downwards is equal to the difference between the weight mg and the reaction R;

Resultant force F= mg-R.

From the second law of motion, the resultant force F=ma.

Therefore, ma=mg-R.

And R=mg-ma =m(g-a).

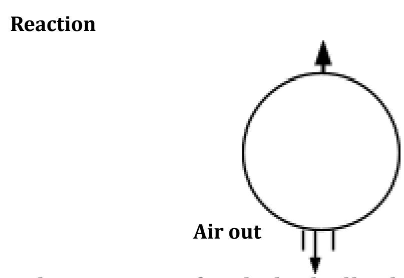
c) When the lift ascends with an acceleration a

In this case, the reaction by the floor of the lift must be greater than the weight of the occupant. Hence, the resultant force F=ma=R-mg.

And R=ma+mg=m(a+g).

The following are some cases where the third law of motion has been applied in everyday life:

- A balloon moves in an opposite direction when air in it is released.



- When a gun is fired, the bullet leaves the gun while the gun recoils backwards.
- For a person running or walking, one exerts a backward force on the ground with the ground exerting a forward push on the foot of the person. This makes running or walking possible.

Example 21.2

1. A man of mass 75kg stands on a weighing machine in a lift. Determine the reading on the weighing machine when the lift:

a) Ascends with an acceleration of $2m/s^2$.

F=ma=R-mg

(75*2)=R-(75*10)

R=150+750 =900N

b) Descends at a constant velocity of 1.5m/s.

$F=ma=mg-R$

But a=0 since the velocity is constant.

Therefore, $75*0=75*10 – R$

$R=750N$

c) Descends with an acceleration of $2.5m/s^2$.

$75*2.5= 75*10 – R$

$R=750 - 187.5=562.5N$

2. A car of mass 1500kg is brought to rest from a velocity of 25m/s by a constant force of 3000N. Determine the change in momentum produced by the force and the time it takes the car to come to rest.

$\Delta P=mv-mu=1500(0-25)=-37500kgm/s$.

$Ft=\Delta P$

We ignore the negative sign in this part because time is a scalar quantity.

$3000*t=37500$

$t=37500/3000 =12.5$seconds.

21.5: Collision and the law of conservation of momentum

This body states that when two or more bodies collide, their total linear momentum before and after collision remain constant provided no external force acts on them;

i.e. momentum before collision= momentum after collision.

There are basically two types of collisions namely elastic and inelastic collision.

a) Elastic collision

This is where the bodies move separate ways after collision. In this collision, not only linear momentum is conserved but also kinetic energy;

- Total linear momentum before collision= total linear momentum after momentum.
- Total kinetic energy before collision= total kinetic energy after collision.

b) Inelastic collision

This is where the colliding bodies stick together and move as one body after collision. In this type of collision, it is only linear momentum which is conserved but not kinetic energy. This is because during this collision, some deformation takes place which eats up part of the energy while some is converted to heat, sound or light energy.

- Total linear momentum before collision= total linear momentum after collision.

Example 21.3

1. A bullet of mass 20g is shot from a gun of mass 20kg with a muzzle velocity of 200m/s. if the bullet is 30cm long, determine:

a) The acceleration of the bullet.

For the bullet: u=0, v=200m/s, s=0.3m

$v^2=u^2+2as$

$200^2=0+(2)(0.3a)$

$a=40000/0.6 =6.667*10^4 m/s^2$

b) The recoil velocity of the gun.

Total linear momentum before collision=total linear momentum after collision

(20*0)+(0.02*0)=(20*v)+(0.02*200)

v=-4/20= -0.2m/s.

2. A 5kg mass moving with a velocity of 10m/s collides with a 10kg mass moving at 7m/s along the same line. If the two masses join together on impact, find their common velocity if they were moving:

a) In opposite directions.

Total linear momentum before collision=total linear momentum after collision

(5*10) + (10*-7)=(5+10)v

15v=-20

v=-20/15 =-1.33m/s

the bodies move in the initial direction of the 10kg mass.

b) In the same direction.

Total linear momentum before collision=total linear momentum after collision

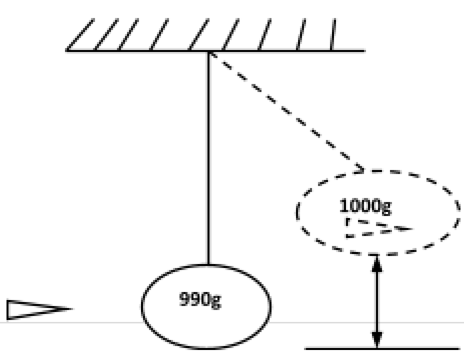
(5*10)+(10*7)=(5+10)v

15v=120

v=120/15 =8m/s

3. A bullet of mass 10g travelling horizontally at 100m/s embeds itself in a block of wood of mass 990g suspended from a light inextensible string so that it can swing freely. Find:

10gh



a) The velocity of the bullet and block immediately after collision.

$(0.01*100)+(0.99*0)=(0.01+0.99)v$

$v=1/1 =1m/s$

b) The height through which the block rises.

At the maximum height, all the kinetic energy is converted into potential energy.

k.e=p.e

$½(mv^2)=mgh$

$½(0.01+0.99)1^2=(0.01+0.99)(10)h$

$h=0.05m$

21.6: Friction

This is a force acting between two surfaces in contact and tends to oppose the intended motion. Friction may be beneficial but can also be a nuisance.

21.6.1: Advantages of friction

- Makes walking, writing possible.
- Required for braking in cars, bicycles etc.
- Makes rotation of the conveyor belts in factories possible.
- Necessary for lighting matchsticks.
- Useful when using nuts, bolts, screw jacks, vices etc.

21.6.2: Limitations of friction

- A lot of energy is lost in the form of heat.

- Causes wear and tear on the pars of machines.

- May lead to noise pollution.

It is therefore important to minimize friction at all cost. This can be done through the following ways:

- Using rollers.
- Using ball bearings.
- Lubrication
- Air cushioning.

21.6.3: Factors affecting friction

Frictional force is directly proportional to the normal reaction R;

$F \alpha R$

Or simply F/R= a constant.

The constant is called coefficient of friction μ. It is a measure of the nature of the surfaces in contact.

Hence, frictional force F= normal reaction R* coefficient of friction μ.

When the two bodies are at rest, then the coefficient of friction is referred to as **coefficient of static** friction while if they are in relative motion, it is called **coefficient of kinetic** friction. Coefficient of friction has no units.

Hence, friction depends on two factors:

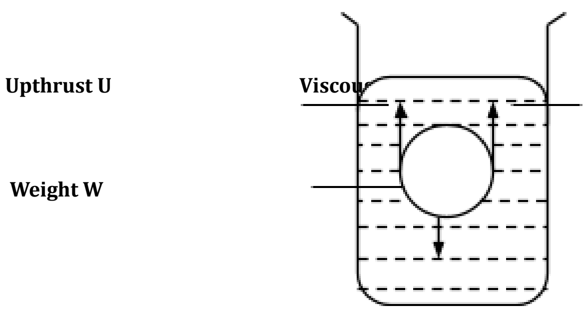
1. The normal reaction R.
2. The nature of the surface. Frictional force is greater between rough surfaces than between smooth surfaces.

Note that frictional force is independent of the area of contact of the two surfaces and relative velocity of the bodies.

21.7: Viscosity

Friction exerted by fluids is called **viscosity** or **viscous drag**. It is the force which opposes relative motion between layers of the fluid. Viscosity is caused by the forces of attraction between the molecules of the fluid. When a body is put in a fluid, three forces act on it, namely:

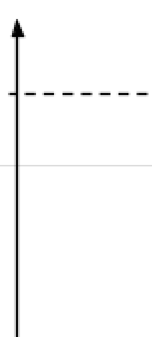
- Weight of the body which acts downwards.
- Upthrust due to the fluid which acts upwards.
- Viscous drag due to the fluid which acts upwards.



When the body enters the fluid, its weight is initially higher than the total upward forces i.e. upthrust plus viscous drag. The resultant force acting on the body accelerates it towards the bottom of the container. As the body sinks down, the viscous drag increases until the three forces balance i.e. W= U+ F. at this point, the body attains its maximum constant velocity called **terminal velocity**. The resultant force on the body is therefore zero.

The graph of velocity against time for a body falling through a fluid appears as shown below:

Terminal velocity



Velocity (m/s)

Time(s)

Note that viscosity decreases with increase in temperature.

TOPIC 22: WORK, ENERGY, POWER AND MACHINES

22.1: Work and Energy

When a force acting on a body displaces the body in the direction of the force **work** is said to have been done. Work is the product of force and displacement in the direction of the force;

Workdone= force F*displacement s.

The SI Unit of work is newton-metre (Nm).

1Nm= 1joule (1J).

A joule is defined as the workdone by a force of one newton to displace a body through one metre in the direction of the force.

Other multiples of the joule include kilojoule(kJ) and megajoule(MJ).

Energy on the hand is the ability or capacity to do work. Anything that possesses energy is capable of doing work. The SI Unit of energy is the joule. Energy has the following characteristics:

- It is not visible.
- Occupies no space.
- Has no mass nor any other physical property.

The most common sources of energy include the sun, wind, geothermal, waterfalls, nuclear or atomic energy, fuels etc.

Energy resources may be grouped into two:

❖ Renewable energy- can be reused again and again. Their supplies are inexhaustible e.g solar, geothermal, wind energy.
❖ Non-renewable energy- their supplies are exhaustible i.e. cannot be reused once exhausted e.g. wood, coal biogas, petroleum etc.

Energy exists in many forms such as mechanical, chemical, heat and electrical energy amongst others. In this topic, we will look at mechanical energy.

22.1.1: Mechanical energy

It is divided into two areas namely kinetic energy and potential energy.

Kinetic energy is the energy possessed by a body in motion. Suppose a body of mass m is moving with a constant velocity v, then its kinetic energy is given by;

Kinetic energy=$\frac{1}{2}(mv^2)$.

Potential energy on the other hand is a form of stored energy in a body when it is in a particular state or position. A body in a raised position possesses **gravitational potential energy** given by;

$P.E_g$=mgh, where m- mass of the body, g- gravitational field strength and h- height above the ground.

Also, a stretched or compressed material is able to regain its original shape when released. This is because it possesses a type of potential energy known as **elastic potential energy**. As can be recalled from Hooke's law, the workdone in stretching or compressing an elastic material is given by;

W=½(Fe) =½(ke^2).

Hence the elastic potential energy is given by;

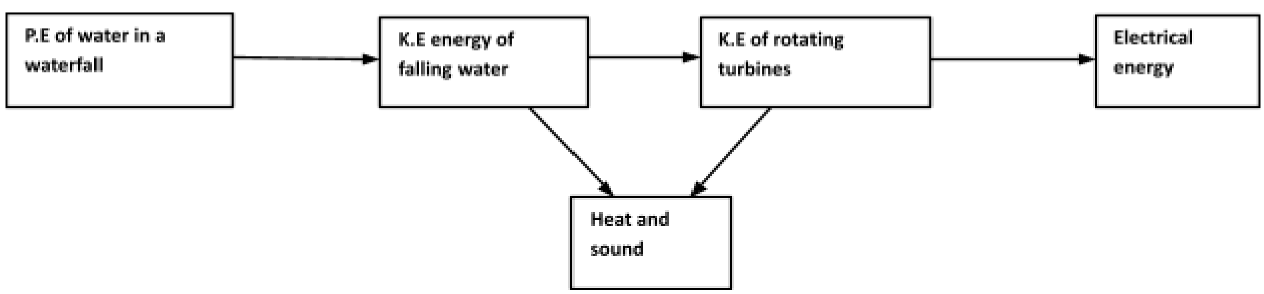
$P.E_e$=½(Fe) =½(ke^2).

22.1.2: The law of conservation of energy

The law states: energy can neither be created nor destroyed but can be transformed from one form to another.

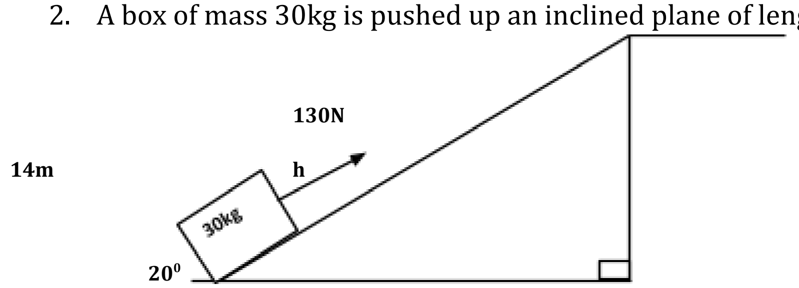
Alternative statement: the sum of kinetic energy and potential energy of a system is a constant.

Below is the energy transformation in a hydroelectric power station:



Example 22.1

1. A force of 40N is applied on a body. The body moves a horizontal distance of 7m. Calculate the workdone on the body.
   W=F*s =40N*7m
   =280Nm or 280J
2. A box of mass 30kg is pushed up an inclined plane of length 14m using a force of 130N as shown below:



If the track is inclined at an angle of 20^0, calculate:

a) The height of the platform.

Sin 20^0=h/14

h=14sin 20^0=

b) Workdone by the force of 130N.

W=F*s = 130*14 =1820J

c) Workdone, if the box is lifted vertically upwards. Compare your answer in (b) and (c) above.

W=mgh=300sin 20^0=

Workdone in pushing the body along the inclined plane is greater than the workdone when lifting the body vertically upwards. This is because of the frictional force between the body and the inclined plane.

d) The frictional force between the box and the inclined plane.

F_r=1820-300sin 20^0=

3. A crane is used to lift a body of mass 30kg through a vertical distance of 6.0m.

a) How much work is done on the body?

W=F*s =(mg)s= 300*6=1800J

b) What is the potential energy stored in the body?

P.E=mgh=30*10*6=1800J

c) Comment on the two answers above.

Workdone on the body is equal to the potential energy stored in the body. Hence the workdone against gravity is stored as the potential energy.

4. A spring of spring constant 25N/m is stretched such that its length increases from 10cm to 20cm. calculate the amount of workdone on stretching the spring.

W=½(ke^2)= ½(25)(0.1^2).

=0.125J

5. A body of mass 12kg is pulled from the rest with a constant force of 25N. The force is applied for 6.0s. Calculate:

a) The distance travelled.

F=ma

a=25N/12kg =2.1m/s^2, u=0, t=6

s=ut+1/2at^2=(0*6)+1/2(2.1)(6^2)=37.8m

b) Workdone on the body.

W=F*s =25*37.8 =945J

c) The final kinetic energy of the body.

K.E=workdone= 945J

d) The final velocity of the body.

K.E=1/2(mv^2)=945J

v={(2*945)/12}$^{1/2}$= 12.6m/s.

22.2: Power

Power is defined as the rate of doing work;

Power=workdone/time.

The SI Unit of power is the watt (W).

1W= 1J/s.

Other multiples of the watt include the kilowatt(kW) and megawatt(MW);

1W=10^{-3}kW

1W=10^{-6}MW

The power of a device is the measure of how fast the device can perform a given task or convert a given amount of energy. For example, a device rated 1kW converts 1000J of energy to another form in one second.

Power=workdone/time =Fd/t.

But d/t =velocity v.

Therefore, power= force F*velocity v.

Example 22.2

1. A person of mass 60kg climbs 16m up a rope in 20seconds. Find the average power developed by the person.

Power=workdone/time =(600*16)/20 =480W

2. A person of mass 40kg runs up a flight of 50stairs each of height 20cm in 5 seconds. Calculate:

a) The workdone.

W=mgh=40*10*(50*0.2)=4000J

b) The average power of the person.

Power=4000J/5s =800W

c) Explain why the energy the person actually uses to climb up is greater than the calculated workdone.

22.3: Machines

A machine is a device that makes work easier. In a machine, a force applied at one point of a system is used to generate another force at a different point of the system to overcome a load. The following terms are used in machines:

a) Effort- the force applied to the machine.

b) **Load-** the force exerted by the machine.

c) **Mechanical advantage (M.A)-** the ratio of the load to effort.

M.A=Load/Effort.

It has no units.

It is dependent on friction between the moving parts and the weight of the parts of the machine that have to be lifted when operating the machine; the greater the friction the smaller the mechanical advantage.

d) **Velocity ratio (V.R)**- it is defined as the ratio of the velocity of the effort to the velocity of the load;

V.R= velocity of effort/velocity of load = $\frac{\text{Effort distance/time}}{\text{Load distance/time}}$

Thus V.R=effort distance/load distance.

Velocity ratio also has no units.

e) **Efficiency η**

It is the ratio of the workdone on the load (work output) to the workdone by the effort (work input) expressed as a percentage;

Efficiency η= (work output/work input)*100.

Efficiency also depends on the friction between the moving parts and the weight of the moveable parts. Hence the efficiency of a machine is always less than 100%.

Efficiency=work output/work input= (load*load distance)/ (effort*effort distance)

= (load/effort)*(load distance/effort distance)

But load/effort =mechanical advantage (M.A),

And, load distance/effort distance =1/velocity ratio

Therefore, efficiency η= (M.A/V.R)*100.

Example 22.3

1. A machine requires 6000J of energy to lift a mass of 55kg through a vertical distance of 8m. Calculate its efficiency.

   Work input= 6000J

   Work output=F*s= 55*10*8 =4400J

   Efficiency=(work output/work input)*100= (4400/6000)*100 = 73.33%

2. An effort of 250N raises a load of 900N through 5m in a machine. If the effort moves through 25m, find:

a) The useful workdone in raising the load.

   Useful workdone=load*load distance =900*5= 4500J

b) The workdone by the effort.

Workdone by the effort= effort*effort distance= 250*25= 6250J

c) The efficiency of the machine.

Efficiency= (work ouput/work input)*100= (4500/6250)*100 = 72%.

3. A machine whose velocity ratio is 8 is used to lift a load of 300N. The effort required is 60N. calculate:

a) The mechanical advantage of the machine.

M.A= load/effort =300/60 =5

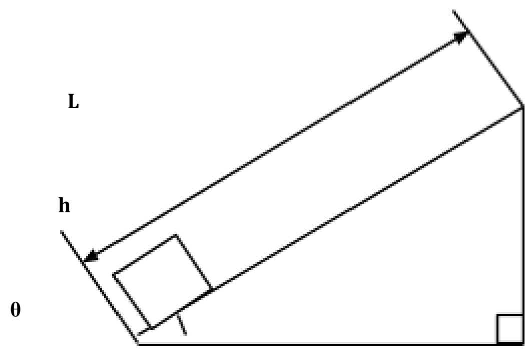
b) The efficiency of the machine.

Efficiency= (M.A/V.R)*100= (5/8)*100 =62.5%

22.4: Types of machines

Below are some of the common machines:

22.4.1: Inclined plane



The distance moved by the effort is L while the vertical height moved by the load is h.

Also, sin θ=h/L

Or simply h=Lsinθ

Therefore, velocity ratio (V.R)= effort distance/ load distance =L/Lsin θ.

Hence V.R= 1/sin θ.

Example 22.4

1. A man uses an inclined plane to lift a 81kg mass through a vertical height of 4.0m. Given that the angle of inclination of the plane is 30^0 and its efficiency is 75%, determine:

a) The effort needed to move the load up the inclined plane at a constant velocity.

V.R=1/sin 30 =2

Therefore, (M.A/2)*100=75

M.A= (2*75)/100 =3/2

3/2 = 810N/effort

Effort= (810*2)/3 =540N

b) The workdone against friction in raising the mass through the height of 4.0m.

Work input=effort*effort distance = (540*4)/sin 30 =4320J

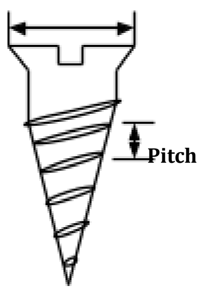
Work output=load*load distance= 81*10*4= 3240J

Therefore, workdone against friction= 4320-3240= 1180J

22.4.2: A screw and bolt

For a screw, when the effort applied on the head moves through a complete revolution, the screw advances by a distance equivalent to one pitch. A pitch is the distance between two successive threads.

d



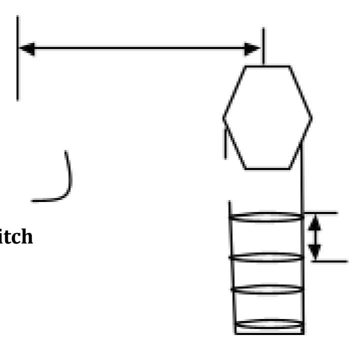
Distance moved by the effort= circumference =πd

Distance moved by the load= one pitch

Hence, velocity ratio (V.R)= circumference/pitch =πd/pitch.

For the bolt, effort is applied at the free end of the spanner.

Radius R



Pitch

Therefore, the distance moved by the effort in one revolution= circumference= 2πR.

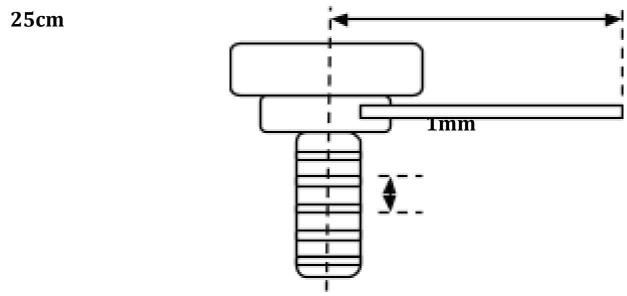
Hence, V.R= circumference/pitch =2πR/pitch.

Note that a combination of a screw and lever can be used as a jack for fitting heavy loads e.g car jack. When two or more systems are combined together, the overall velocity ratio is the product of the individual velocity ratios;

Combined V.R= $V.R_1 * V.R_2 * \ldots\ldots\ldots\ldots * V.R_k$

Example 22.5

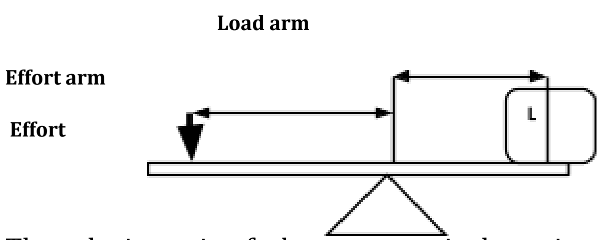
1. The figure below shows a screw jack whose screw has a pitch of 1mm and has a handle of 25cm long.



Determine the velocity ratio of the jack.

V.R= $2\pi r$/pitch= 2π(25cm)/0.1cm= 1571

22.4.3: Lever system



The velocity ratio of a lever system is the ratio of the effort arm to the load arm;

V.R= Effort arm/ Load arm.

22.4.4: Gears

A gear is a wheel with equally spaced teeth or cogs around it. The wheel on which the effort is applied is called the driving (input) gear while the load gear is referred to as the driven (output) gear. Suppose the driving gear has **n** teeth and the driven gear**N** teeth, then when the driving gear makes one complete revolution the driven gear makes **n/N** revolutions.

V.R of the system = Number of revolutions made by the effort (driving) gear

Number of revolutions made by the load (driven) gear.

V.R = 1revolution =N/n

n/N revolutions

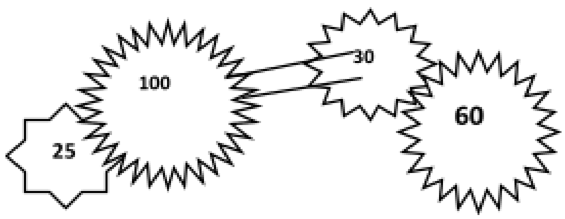
Hence, velocity ratio of a gear system is the ratio of the number of teeth of the driven gear to the number of teeth of the driving gear;

$$V.R = \frac{\text{Number of teeth of the driven gear}}{\text{Number of teeth of the driving gear}}$$

Example 22.6

1. A driving gear having 25teeth engages with a second gear with 100teeth. A third gear with 30 teeth on the same shaft as the second one engages with a fourth gear having 60teeth. Find:

a) The total velocity ratio of the system.



Combined $V.R = V.R_1 * V.R_2$

$V.R_1$=No. of teeth of driven gear/ No. of teeth of driving gear

= 100/25= 4

$V.R_2$=60/30= 2

Hence, V.R= 4*2= 8

b) The mechanical advantage of the system if its efficiency is 85%.

Efficiency= (M.A/8)*100= 85

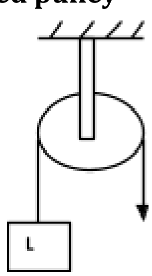
M.A= (85*8)/100= 6.8

22.4.5: Pulleys

A pulley is a wheel with a groove to accommodate a string or rope. There are three possible systems of pulleys namely single fixed, single moveable and a block and tackle.

a) Single fixed pulley

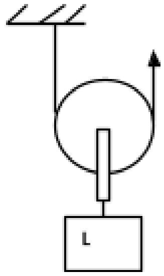
E



In this arrangement, both the effort and load move through the same distance. Hence the velocity ratio of the system is one.

b) Single moveable pulley

E



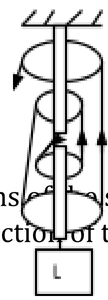
The load is supported by two sections of the string. If the load is pulled upwards through a distance of 1m, each section of the string also moves through 1m. Hence the effort moves through a total distance of 2m.

Therefore, the velocity ratio of the system = effort distance/load distance =2m/1m =2.

c) A block and tackle

This system comprises two sets; one set fixed and the other moveable. A single string is then passed around each pulley in turn. The arrangement can take several forms depending on the desired velocity ratio. Below is an example:

E



In this case, there are four sections of the string supporting the load. Hence, when the load moves upwards through a distance of 1m, each section of the string also shortens by 1m. Therefore, the total distance moved by the effort (string) is 4m.

Thus, V.R of the system= effort distance/load distance =4m/1m =4. Coincidentally, the velocity ratio of the system is the same as the number of sections of the string supporting the load.

Generally, the velocity ratio of a block and tackle system is given by the number of sections of the string supporting the load.

Practically, the efficiency of any pulley system is less than 100%. This is as a result of two reasons:

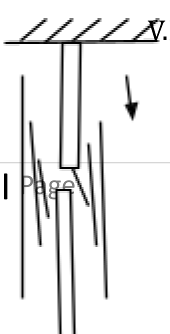
- The friction between the moveable parts.
- The weight of the parts that have to be lifted when operating the system.

Example 22.7

1. The figure below shows a pulley system used to raise a load.

    a) State the velocity ratio of the system.

    V.R=number of strings supporting the load= 6



b) If an effort of 1000N is needed to raise a load of 4500N, determine the efficiency of the system.

M.A= load/effort= 4500N/1000N= 4.5

Efficiency = (M.A/V.R)*100= (4.5/6)*100= 75%

c) Calculate the wasted energy if a mass of 500kg is lifted up through a height of 2m using the same system.

Work ouput= load*load distance= 500*10*2= 10000J

Efficiency= (work output/work input)*100

Therefore, (10000J/work input)*100=75

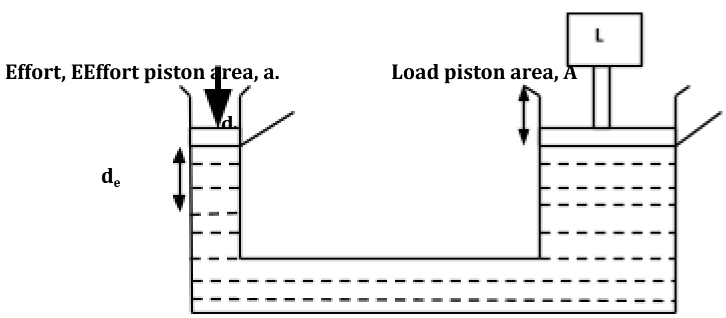
Work input= (10000*100)/75 =13333.33J

Wasted energy= 13333.33-10000= 3333.33J

Alternatively, wasted energy=25% of work input= (25/100)*13333.33J=3333.33

22.6: Hydraulic machine

Consider the diagram below:



When the effort is applied as shown, the volume of the liquid leaving the effort arm is the same as the volume of the liquid entering the load arm;

i.e. $a*d_e=A*d_l$,

$\mathbf{d_e/d_l= A/a}$

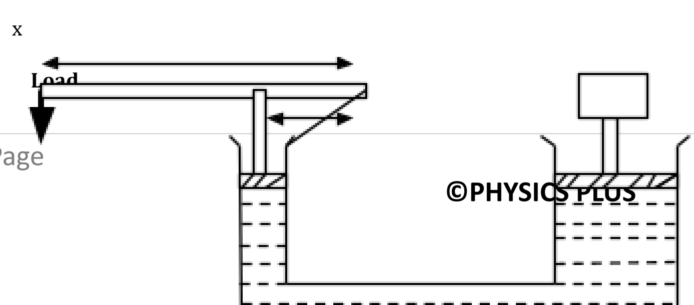
Therefore, the velocity ratio of a hydraulic system is the ratio of the area of the load piston to the area of the effort piston. If the pistons are circular then;

V.R=area of load piston/area of effort piston $=\pi R^2/\pi r^2$

V. $=R^2/r^2$, where R- is the radius of the load piston and r- is the radius of the effort piston.

Example 22.8

1. In the figure below x=30cm, y=6cm, effort E=60N, A_1=4cm^2 and A_2=12cm^2.



Ey

A_1A_2

Calculate:

a) The force F exerted on the liquid at A_1.

By the principle of moments;

60N*30cm= F*6cm

F= (60*30)/6= 300N

b) The velocity ratio of the system.

V.R of the lever system= effort arm/load arm =30cm/6cm= 5

V.R of the hydraulic system= area of load piston/area of effort piston= $12cm^2/4cm^2$= 3

Therefore, the combined V.R= 5*3= 15

c) The maximum load that can be raised by the system.

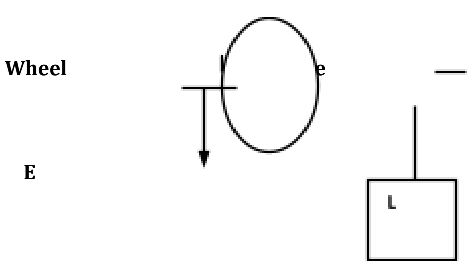
Pressure at A_1= Pressure at A_2

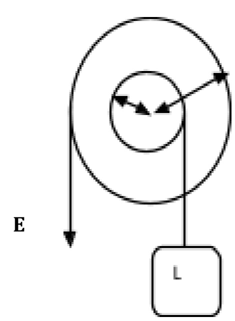
$300N/4cm^2 = L/12cm^2$

L= (300*12)/4 =900N.

22.7: Wheel and axle

It consists of a large wheel of radius R attached to an axle of radius r.





Note that in this case, both the wheel and axle make the same number of revolutions at any time;

Thus, in one revolution the distance moved by the effort= $2\pi R$,

And the distance moved by the load= $2\pi r$.

Hence, the velocity ratio of the system= $2\pi R/2\pi r = R/r$.

Thus the velocity ratio of a wheel and axle is the ratio of the radius of the wheel to the radius of the axle.

Example 22.9

1. A wheel and axle is used to raise a load of 140N by a force of 20N applied to the brim of the wheel. If the radii of the wheel and axle are 70cm and 5cm respectively, calculate the mechanical advantage, velocity ratio and efficiency of the system.

   M.A= load/effort =140N/20N= 7

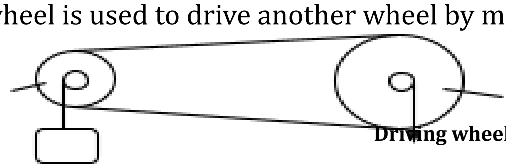
   V.R =radius of the wheel/radius of the axle= 70cm/5cm= 14

   Efficiency= (M.A/V.R)*100

   = (7/14)*100= 50%

22.8: Pulley belt

This is where one wheel is used to drive another wheel by means of a belt.



The driving wheel covers a distance 2πR in one revolution while the driven wheel covers a distance 2πr in one revolution. If the driving wheel makes one revolution, the driven wheel makes 2πR/2πr (R/r) revolutions.

V.R of the system= Number of revolutions made by the effort (driving) wheel

Number of revolutions made by the load (driven) wheel

V.R = 1/(R/r) =r/R

Therefore, the velocity ratio of a pulley belt is the ratio of the radius of the driven (load) wheel to the radius of the driving (effort) wheel.

TOPIC 23: HEATING EFFECT OF ELECTRIC CUURENT

23.1: Introduction

When current flows through a conductor, heat energy is generated in the conductor. The heating effect of an electric current depends on three factors:

- The resistance, R of the conductor. A higher resistance produces more heat.
- The time, t for which current flows. The longer the time the larger the amount of heat produced
- The amount of current, I. the higher the current the larger the amount of heat generated.

Hence the heating effect produced by an electric current, I through a conductor of resistance, R for a time, t is given by $H = I^2Rt$. This equation is called the Joule's equation of electrical heating.

23.2: Electrical energy and power

The work done in pushing a charge round an electrical circuit is given by w.d = VIt

So that power, $P = w.d / t = VI$

The electrical power consumed by an electrical appliance is given by $P = VI = I^2R = V^2/R$

Example 23.1

1. An electrical bulb is labeled 100W, 240V. Calculate:

a) The current through the filament when the bulb works normally

b) The resistance of the filament used in the bulb.

{ **ans. 0.4167A, 576.04Ω**}

Solution

a) $I = P/V = 100/240 = 0.4167A$

b) $R = P/I^2 = 100/ 0.4167^2 = 576.04\Omega$ or $R = V^2/P = 240^2/100 = 576\Omega$

2. Find the energy dissipated in 5 minutes by an electric bulb with a filament of resistance of 500Ω connected to a 240V supply. { **ans. 34,560J**}

Solution

$E = Pt = V2/R * t = (240^2 * 5 * 60)/500 = 34{,}560J$

3. A 2.5 kW immersion heater is used to heat water. Calculate:

   a) The operating voltage of the heater if its resistance is 24Ω

   b) The electrical energy converted to heat energy in 2 hours.

   {**ans. 244.9488V, 1.8*10⁷J**}

Solution

a) $P=VI=I^2R$

$I = (2500/24)^{1/2} = 10.2062A$

$V=IR= 10.2062 * 24 = 244.9488V$

b) $E = VIt = Pt = 2500*2*60*60 = 1.8 * 10^7J$

OR $E= VIt = 244.9488 * 10.2062 * 2 * 60 * 60 = 1.8 * 10^7J$

4. An electric bulb is labeled 100W, 240V. Calculate:

a) The current through the filament

b) The resistance of the filament used in the bulb.

{**ans. 0.4167A, 575.95Ω**}

Solution

a) $P = VI \quad I = P/V = 100/240 = 0.4167A$

b) From Ohm's law, V =IR   R=V/I =240/0.4167 = 575.95Ω

23.3: Applications of heating effect of electric current

Most household electrical appliances convert electrical energy into heat by this means. These include filament lamps, electric heater, electric iron, electric kettle, etc.

In lighting appliances

a) Filament lamps- it is made of a tungsten wire enclosed in a glass bulb from which air has been removed. This is because air would oxidize the filament. The filament is heated up to a high temperature and becomes white hot. Tungsten is used due its high melting point; 3400^0C. The bulb is filled with an inactive gas e.g. argon or nitrogen at low pressure which reduces evaporation of the tungsten wire. However, one disadvantage of the inert gas is that it causes convection currents which cool the filament. This problem is minimized by coiling the wire so that it occupies a smaller area which reduces heat loss through convection.

b) Fluorescent lamps- these lamps are more efficient compared to filament lamps and last much longer. They have mercury vapour in the glass tube which emits ultraviolet radiation when switched on. This radiation causes the powder in the tube to glow (fluoresce) i.e. emits visible light. Different powders produce different colours. Note that fluorescent lamps are expensive to install but their running cost is much less.

In electrical heating

c) Electric cookers- electric cookers turn red hot and the heat energy produced is absorbed by the cooking pot through conduction.

d) Electric heaters- radiant heaters turn red at about 900^0C and the radiation emitted is directed into the room by polished reflectors.

e) Electric kettles- the heating element is placed at the bottom of the kettle so that the liquid being heated covers it. The heat is then absorbed by water and distributed throughout the whole liquid by convection.

f) Electric irons- when current flows through the heating element, the heat energy developed is conducted to the heavy metal base raising its temperature. This energy is then used to press clothes. The temperature of the electric iron can be controlled using a thermostat (a bimetallic strip).

TOPIC 24: QUANTITY OF HEAT

24.1: Introduction

When heat is transferred from one body to another, the body which loses heat has its temperature lowered while that which gains heat has its temperature raised.

24.2: Terms used

Heat capacity, C.

This is the quantity of heat energy required to raise the temperature of a given mass of substance by one Kelvin.

i.e. heat capacity, **$C = Q\ (J)/\Delta\theta\ (K)$**

Hence the SI Unit of heat capacity is joule per Kelvin (JK^{-1}).

Specific heat capacity, c

This is the quantity of heat energy required to raise the temperature of a unit mass of a substance by one Kelvin. i.e**. c = Q (J)/mΔθ (KgK)**

Q = mcΔθ

The SI Unit of specific heat capacity is joules per kilogram per Kelvin ($JKg^{-1}K^{-1}$).

Note that **c = C/m**

Therefore heat capacity, C = mass, m * specific heat capacity, c.

The table below shows some substances with their specific heat capacities:

Material	s.h.c ($JKg^{-1}K^{-1}$)
Water	4200
Alcohol	2300
Kerosene	2200
Ice	2100
Aluminium	900
Glass	830
Iron	460
Copper	390
Mercury	140
Lead	130

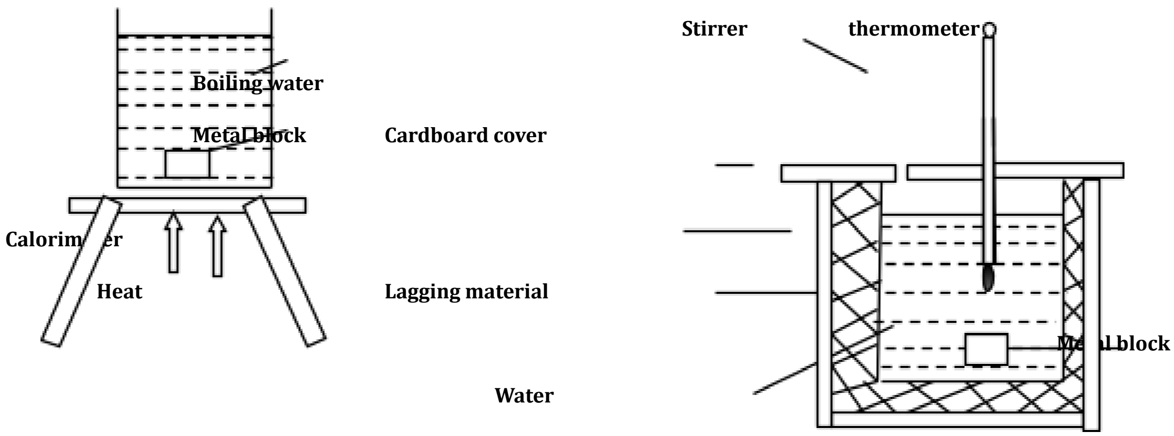
24.2.1:

Determination of the specific heat capacity

By the method of mixtures

a) **S.h.c of solids**

In this method, a known mass of a solid, e.g. a metal block is heated by dipping it in a bath of hot water. After some time, the solid is very fast transferred into cold water in a calorimeter and whose mass is known.



The calorimeter is then covered using a piece of cardboard and stirred continuously. The following measurements are then recorded:

- Mass of the solid metal block, m_s
- Mass of copper calorimeter with the stirrer, m_c
- Mass of the calorimeter and stirrer + water, m_1
- Temperature of the boiling water (initial temperature of the metal block), θ_s
- Temperature of cold water in the calorimeter (initial temperature of calorimeter), θ_w
- Final steady temperature of the mixture, θ

Calculation

Mass of the water in the calorimeter = $m_1 - m_c = m_w$

Temperature change of the hot metal block = $\theta_s - \theta$

Temperature change of the water in the calorimeter and the calorimeter = $\theta - \theta_w$

Assuming there is no heat loss to the surrounding when the metal block is being transferred into the cold water and thereafter;

Amount of heat lost by the metal block = amount of heat gained by calorimeter with stirrer + amount of heat gained by water in the calorimeter.

i.e. $m_s c_s(\theta_s - \theta) = m_c c_c(\theta - \theta_w) + m_w c_w(\theta - \theta_w)$

wherec_s – s.h.c. of the metal block

c_c – s.h.c. of the copper calorimeter

c_w – s.h.c. of water.

Hence s.h.c. of the metal block, $c_s=[m_c c_c(\theta-\theta w) + m_w c_w(\theta-\theta_w)] / m_s(\theta_s-\theta)$

b) S.h.c. of a liquid

In this case, a solid of known s.h.c. is used and the water in the calorimeter is replaced with the liquid whose s.h.c. is to be determined. The solid metallic block is first heated in a bath of boiling water and then transferred into the calorimeter containing the liquid. The following measurements are then collected:

- Mass of the metal block, m_s
- Mass of the calorimeter with stirrer, m_c
- Mass of the calorimeter, stirrer and the liquid, m_1
- Initial temperature of the metal block, θ_s
- Initial temperature of the liquid, θ_l
- Final steady temperature of the mixture, θ

If the there is no heat loss to the surrounding, then the quantity of heat lost by the metal block equals the quantity of heat gained by the calorimeter with stirrer and the liquid.

i.e. $m_s c_s(\theta_s – \theta) = [m_c c_c(\theta-\theta_l) + m_l c_l(\theta-\theta_l)]$

Hence $c_l = [m_c c_s(\theta_s – \theta) - m_c c_c(\theta-\theta_l)] / m_l(\theta-\theta_l)$

Alternatively the s.h.c. of a liquid can be obtained by mixing it with another liquid whose specific heat capacity is known and their common temperature determined.

The following precautions must be taken to minimize heat losses to the surroundings:

- Using a highly polished calorimeter
- Heavily lag the calorimeter
- Using a lid of poor thermal conductivity

Example 24.1

1. 70g of a solid initially at 25^0C was carefully dropped into water in a calorimeter at 60^0C. If the final constant temperature of the water and the solid was 54^0C and the mass of water is 500g, determine the specific heat capacity of the solid. Assume the heat absorbed by the calorimeter to be negligible. Take the s.h.c. of water = $4200JKg^{-1}K^{-1.}$

   {**ans. 10, 769.23JKg^{-1}K^{-1}**}

   Solution

   Heat lost = heat gained

$m_w c_w \Delta\theta_w = m_s c_s \Delta\theta_s$

0.5Kg * 4200JKg^{-1}K^{-1} * (60-54) K = 0.07kg * c_s * (54-25) K

Cs = 29400J / 2.73KgK = 1O, 769.23 JKg^{-1}K^{-1}

2. A student heated 20Kg of water to a temperature of 80^{0}C. He then added x Kg of water at 15^{0}C and the final steady temperature of the mixture is 40^{0}C. Given that the s.h.c. of water is 4.2Jg^{-}1K^{-1}, determine the value of x. {**ans. 32kg**]

   Solution

Heat lost = heat gained

20kg * 4200Kg^{-1}K^{-1} * (80-40) K = x * 4200JKg^{-1}K^{-1} * (40-15) K

X = 3, 360, 000/105, 000 = 32kg.

3. 0.2kg of iron at 100^{0}C is dropped into 0.09kg of water at 26^{0}C inside a calorimeter of mass 0.15kg and s.h.c. 800JKg^{-1}K^{-1}. Find the final temperature of the water. Take the s.h.c. of iron = 460JKg^{-1}K^{-1} and that of water = 4200JKg^{-1}K^{-1}.

   {**ans. 37.2^{0}C**}

   Solution

Heat lost by iron = heat gained by calorimeter + heat gained by water.

0.2kg * 460JKg^{-1}K^{-1} * (100-θ_c) K = 0.15kg * 800JKg^{-1}K^{-1} * (θ_C-26) + 0.09Kg * 4200JKg^{-1}K^{-1} * (θ_c-26)

9200-92θ_c = 126θ_c-3120 + 378θ_c-9828

596θ_c = 22148

θ_c = 22148 / 596 = 37.2^{0}C

4. A certain block is heated such that its temperature is raised from 15^{0}C to 45^{0}C. calculate the amount of heat absorbed by the metal if its heat capacity is 460JK^{-1} {**13, 800J** }

   Solution

Q = C * $\Delta\theta$ =460JK^{-1} * (45-15) K = 13, 800J.

5. In an experiment to determine the specific heat capacity of a metal, a 100g of the metal was transferred from boiling water to a lagged copper calorimeter containing cold water. The water was stirred and a final steady temperature was realized. The following data was recorded:

   -initial temperature of cold water and calorimeter =20^{0}C

   -temperature of boiling water =99^{0}C

   -final temperature of water, calorimeter and metal =27.7^{0}C

   -mass of cold water plus calorimeter =130g

   -mass of calorimeter =50g

Take s.h.c. of water=4200JKg-1K-1, s.h.c. of copper=400JKg-1K-1.

Use the data above to determine:

a) The heat gained by the water and calorimeter

$Q = mc\Delta\theta_w + mc\Delta\theta_{c\,=}$ (0.08*4200*7.7) + (0.05*400*7.7)

= 2741.2J

b) The specific heat capacity of the metal

0.1*c*71.3=2741.2

C=2741.2/0.1*71.3 = 384.46$JKg^{-1}K^{-1}$

c) State the possible sources of error in the value of the s.h.c obtained in the above experiment.

- Heat loss as the metal was being transferred from the boiling water to the calorimeter.
- Error when reading the thermometer (parallax error)

6. 3kg of hot water was added to 9kg of cold water at 10^0C and the resulting temperature was 20^0C.ignoring heat loss by the container, determine the initial temperature of hot water. Take s.h.c of water=4200$JKg^{-1}K^{-1}$.

$mc\Delta\theta_h = mc\Delta\theta_c$

$3*(\theta-20) = 9*10$

$3\theta=90+60 =150$

$\theta = 150/3 = 50^0C$

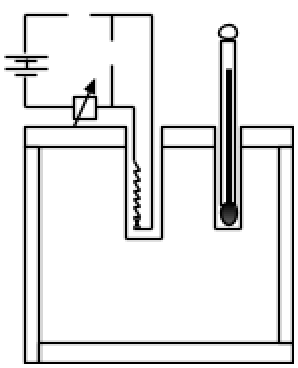
Electrical method

a) Specific heat capacity of a solid

In this method, two holes are drilled in the solid to accommodate the heater and thermometer. The solid is heated electrically for a given time. Below is an arrangement that can be used:

A

V



In this method, the following data is recorded:

- Mass of the metal (solid)
- Heater voltage, V
- Heater current, I
- Time (duration) of heating, t
- Initial temperature of the solid
- Final temperature of the solid

The electrical energy lost by the heater is given by; E = VIt

Suppose there is no heat loss to the surroundings, then the heat lost by the heater equal heat gained the solid.

i.e. $VIt = mc\Delta\theta$

Hence $c = VIt/m\Delta\theta$

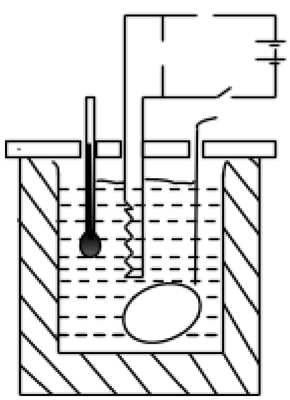
Note

Heat loss is minimized by lagging the calorimeter as well as oiling the holes.

Specific heat capacity of a liquid

A

V



The heat lost by the heater equal the heat gained by the liquid and the calorimeter.

$VIt = mc\Delta\theta_l + mc\Delta\theta_c$

Hence $c_l = (VIt - mc\Delta\theta_c)/m\Delta\theta_l$

Example 24.2

1. An immersion heater rated 120W, 240V is connected to a 240V power supply. How long will it take to heat 1 kg of water from 10^0C to 90^0C? Take s.h.c of water=$4200JKg^{-1}K^{-1}$.

$t= mc\Delta\theta/VI =mc\Delta\theta/P$

$t= (1*4200*80)/120 =2800$seconds.

2. A heater rated 180W and a thermometer were inserted in a 0.5kg of water in a copper calorimeter. The following results were recorded:

Temperature,T(^{0}C)	30	36	40	45	49	54	57
Time,t(minutes)	3	4	5	6	7	8	9

a) Plot a graph of temperature against time

b) Use the graph to find:

- The room temperature
- The specific heat capacity of water.

3. A 180W heater is immersed in a copper calorimeter of mass 100g containing 200g of alcohol. When the heater is switched on, after 36 seconds the temperature of the calorimeter and its contents was raised by 120C. Find the specific heat capacity of alcohol. Take the s.h.c of copper=400$JKg^{-1}K^{-1}$.

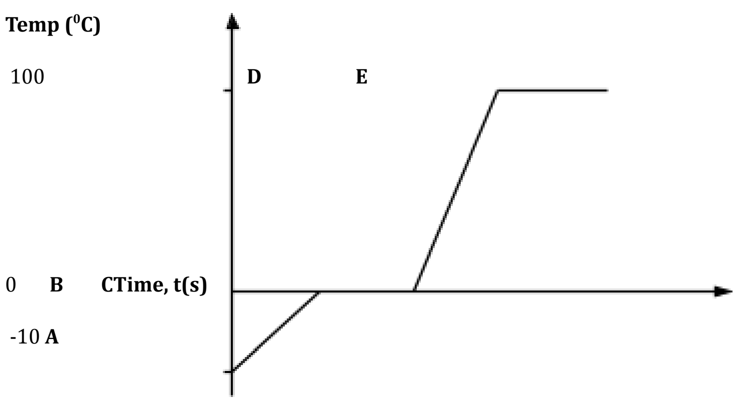
$Pt=mc\Delta\theta_a + mc\Delta\theta_c$

$C_a= (pt-mc\Delta\theta_c)/m\Delta\theta_a= (180*36-0.1*400*12)/0.2*12$

$= 2500JKg^{-1}K^{-1}$

24.3: CHANGE OF STATE

When ice is heated say from -10^{0}C until it boils, it undergoes changes which can be represented by the heating curve below:



Between the points AB, ice absorbs heat energy and its temperature rises. Between BC, the ice absorbs its latent heat of fusion which it uses to melt. This change of state occurs at a constant temperature. Between CD water absorbs heat energy as its temperature rises until boiling point. As the water boils at constant temperature, it absorbs its latent heat of vaporization.

When the vapour condenses to liquid, it gives out its latent heat of vaporization. Similarly, when a liquid freezes to solid, it gives out its latent heat of fusion.

Note:

Latent heat of fusion- it is the quantity of heat needed to convert a given mass of a solid to liquid at constant temperature.

Specific latent heat of fusion- it is the quantity of heat needed to convert a unit mass of a solid to liquid at constant temperature. i.e. $\mathbf{l_f = Q/m}$

Therefore $\mathbf{Q=ml_f}$

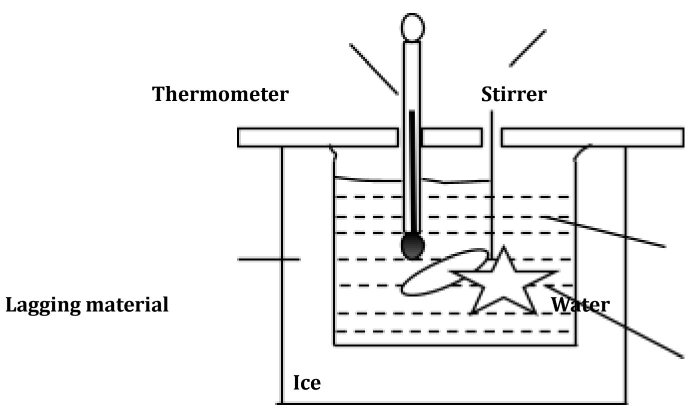
The SI unit of the specific latent heat of fusion is the joule per kilogram (JKg^{-1}). A unit mass of a substance changing from liquid to solid will give out heat energy equivalent to its specific latent heat of fusion.

24.3.1: Determination of specific latent heat of fusion.

There are two methods used:

Mixture method

A piece of dry ice is dropped into a calorimeter containing water slightly above room temperature. Stir the mixture until all the ice has melted. Suppose there is no heat loss to the surroundings, then the heat energy lost by the water and calorimeter equals the heat energy gained by the melting ice



In the above experiment, the following data is recorded for purposes of determining the specific latent heat of fusion:

-Mass of the dry ice

-mass of the water in the calorimeter

-mass of the calorimeter plus stirrer

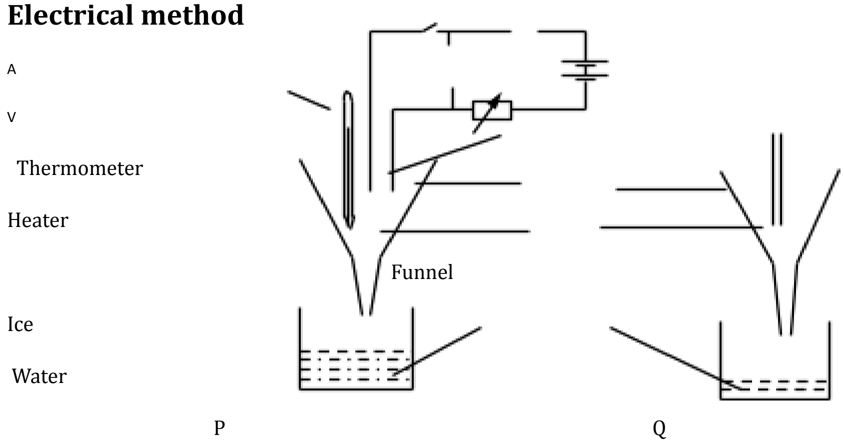
-Temperature change of the water

Hence $mc\Delta\theta_w + mc\Delta\theta_c = ml_f$

$L_f = (mc\Delta\theta_w + mc\Delta\theta_c)/m_i$

NoteDry ice is used due to its low moisture content. This implies that all the heat absorbed by the ice is used to melt the ice and not warming the moisture.

Electrical method



Equal amounts of crushed ice are put simultaneously in two identical filter funnels. A heater is then immersed in the funnel in set up **P**. Place clean dry beakers below each funnel. Wait until a reasonable amount of water has collected in the beaker **P** then switch off the heater and remove the beakers. Weigh the beakers and their contents.

In the above experiment, the following data is collected:

- Mass of the beaker under **P** before experiment, m_1
- Mass of the beaker under **P** after the experiment, m_2
- Mass of beaker under **Q** before experiment, m_3
- Mass of beaker under **Q** after experiment, m_4
- Heater voltage, V
- Heater current, I
- Duration of heating, t

Calculations

Mass of melted ice in set up **P,**$m_p = m_2 - m_1$

Mass of melted ice in set up **Q**, $m_q = m_4 - m_3$

Set up **Q** is called the control experiment. It helps to determine the mass of ice that melted as a result of the temperature of the room during the experiment. In order to obtain the mass of ice melted by the heater only, it is important to subtract the mass of melted ice in Q from that melted in P;

i.e. $m = m_p - m_q$.

Then, heat energy supplied by the heater = heat energy absorbed by the melting ice.

$VIt = ml_f$

Hence $l_f = VIt/m$

The table below gives some common solids and their specific latent heats of fusion:

Material	**s.l.h of fusion (*10^5) JKg^{-1}**
Copper	4.0
Aluminium	3.9
Water(ice)	3.34
Iron	2.7
Wax	1.8
Naphthalene	1.5
Solder	0.7
Lead	0.026
Mercury	0.013

Example 24.3

1. A block of ice of mass 40g at 0^0C is placed in a calorimeter containing 400g of water at 20^0C. Ignoring heat absorbed by the calorimeter, determine the final temperature of the mixture after all the ice has melted. Take s.h.c.of water= $4200JKg^{-1}K^{-1}$ and the s.l.h. of fusion of ice= 340, $000JKg^{-1}$.

   Heat lost by the hot water= heat gained by melting ice + heat gained by melted ice

   $mc\Delta\theta_h = ml_f + mc\Delta\theta_m$

   $0.4*4200*(20-\theta) = (0.04*340, 000) + (0.04*4200*\theta)$

   $33600-1680\theta = 13600 + 168\theta$

   $1848\theta=20000$

   $\Theta=20000/1848 =10.82^0C$

2. 16g of dry ice was added to 100g of water at 26^0C in a beaker of negligible heat capacity. After the ice had all melted, the temperature of water was found to be 11^0C. Find the specific latent heat of fusion of ice. Take the s.h.c of water =$4200JKg^{-1}K^{-1}$.

   $0.1*4200*(26-11) = (0.016*l_f) + (0.016*4200*11)$

   $6300=0.016l_f + 739.2$

   $L_f=5560.8/0.016 = 3.4755 * 105JKg^{-1}$

3. An aluminium tray of mass 400g containing 300g of water is placed in a refrigerator. After 80minutes, the tray is removed and it is found that 60g of water remains unfrozen at 0^0C. If the initial temperature of the tray and its contents was 20^0C, determine the average amount of heat removed per minute by the refrigerator. Take s.h.c of aluminium = $900JKg^{-1}K^{-1}$, s.h.c of water = 4200JKg-1K-1, s.l.h. of fusion of ice = $3.4*105JKg^{-1}$.

Heat lost by tray = $mc\Delta\theta$ = 0.4*900*(20-0)= 7200J

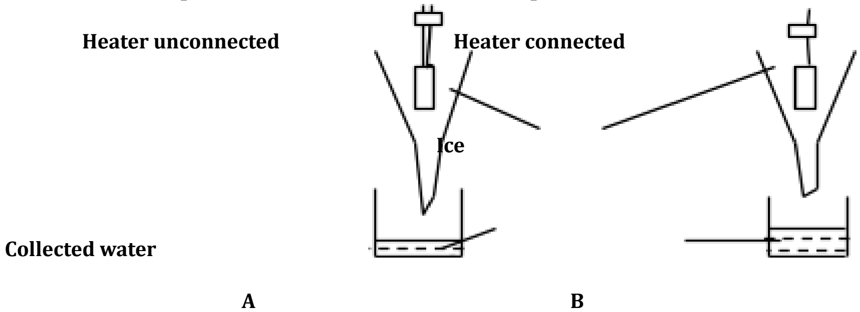
Heat lost by water = $mc\Delta\theta$= 0.3*4200*20 = 25, 200J

Latent heat of ice given out = ml_f = (0.3-0.06)*340, 000 = 999, 600J

Total heat energy absorbed by the refrigerator =3600+25200+999600= 114000J

Hence amount of heat removed per minute = 114000J/80min = 1425J/min

4. In an experiment to determine the specific latent heat of fusion of ice, the following set up was used:



In A the heater is unconnected and when the ice is melting steadily, 0.015kg of water is collected in 300s. In B the heater is connected to a power supply rated 50W. When water drips at a steady rate, 0.058kg of water is collected in 300s. Calculate the value for the specific latent heat of fusion of ice.

Q= Pt = ml_f

L_f= (50*300)/(0.058-0.015)

= 348, 837.21JKg^{-1}

Latent heat of vaporization

This is the quantity of heat energy required to convert a given mass of a liquid to gas at constant temperature.

Specific latent heat of vaporization

This is the quantity of heat energy required to convert a unit mass of a liquid to gas at constant temperature.

L_v = Q/m

Therefore, Q= ml_v

The SI unit of specific latent heat of vaporization is the joule per kilogram (JKg^{-1}).

24.3.2: Determination of the specific latent heat of vaporization

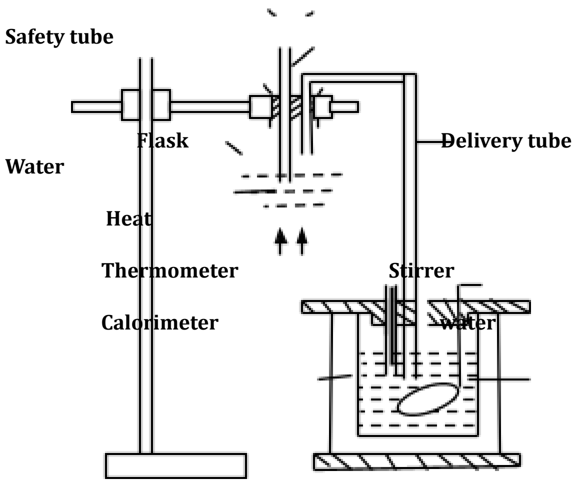
Experiment

Aim: To determine the specific latent heat of vaporization of water using mixture method**.**

Apparatus

- Calorimeter with a stirrer
- Water

- Thermometer
- Flask with a delivery tube
- Heat source
- Weighing machin



Procedure

1. Set up the apparatus as shown above.
2. Find the mass of the calorimeter when empty and when filled with water to the level shown.
3. Measure and record the initial temperature of water in the calorimeter.
4. Heat the water in the flask until it delivers steam through the delivery tube. Ensure that the free end of the delivery tube is inside water in the calorimeter.
5. Allow steam to bubble into the water while stirring until the temperature of water rises by about 200C above the room temperature.
6. Remove the delivery tube from the calorimeter and record the temperature of the water.
7. Determine the new mass of the calorimeter and its contents. Hence, determine the mass of the condensed steam.

Note

Steam first condenses to water which then cools down, losing heat energy.

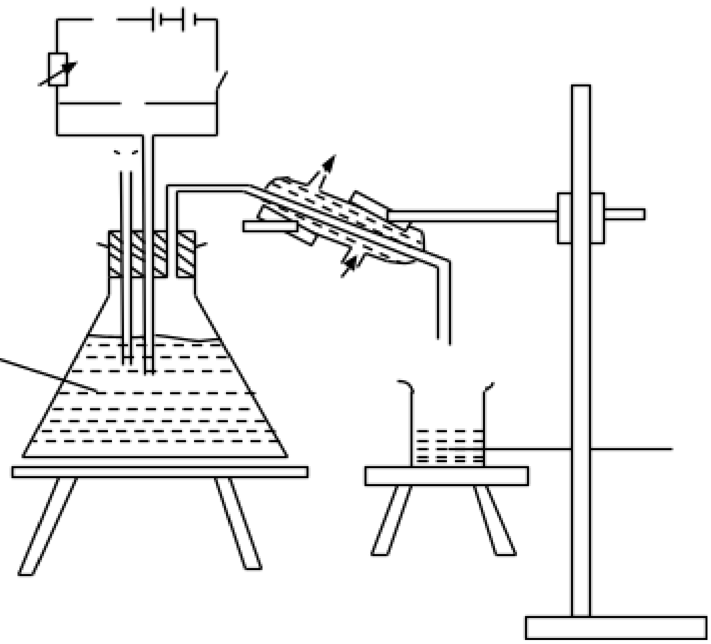
Therefore, heat energy lost by steam and the cooling water equal to the heat energy gained by the water and calorimeter.

$ml_v + mc\Delta\theta_h = mc\Delta\theta_w + mc\Delta\theta_c$

$L_v = (mc\Delta\theta w + mc\Delta\theta_c - mc\Delta\theta_h)/m$

It is important to first cool the water in the calorimeter to a certain value below the room temperature and then pass the steam through it until the temperature rises above the room temperature by the same value. This will help minimize errors due to the heat loss to the surrounding.

Specific latent heat of vaporization using the electrical method



The heating process is allowed to continue until a steady state where condensed water drips out at a constant rate has been achieved. The mass of water collected after a time, t is measured. The following data is collected in this experiment:

- Heater current, I

- Heater voltage, V

- Mass of empty beaker

- Mass of beaker and collected water

- Time taken to collect the condensed water

Suppose all the heat given by the heater is used to convert water to steam, then:

$VIt = ml_v$

Hence, $l_v = VIt/m$

The table below shows some common liquids and their specific latent heats of vaporization;

Liquid	**s.l.h. of vaporization * 10^5(JKg-1)**
Water	22.6
Alcohol	8.6

Ethanol	8.5
Petrol	6.3
Benzene	4.0
Ether	3.5
Turpentine	2.7

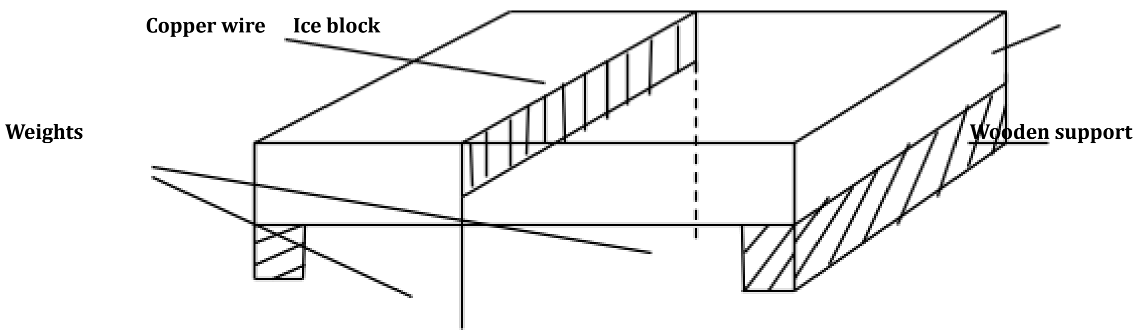
24.4: Boiling and Melting

Boiling and melting points are generally affected by two factors; impurities and pressure.

Melting

1.Effects of pressure on the melting point

Increase in pressure lowers the melting point of a material. This can be illustrated by suspending two weights supported by a copper wire on the surface of an ice block as shown below:



The wire is seen to cut its way through the block of ice but leaves it as one piece. The suspended weights make the copper wire to exert pressure on the ice directly underneath which is made to melt at a temperature below its melting point. As the wire cuts through, the water formed flows over the wire and immediately solidifies since it is no longer under pressure. As the water solidifies, it gives out its latent heat of fusion which is conducted by the copper wire to melt the ice below it. This continues until the copper wire completely cuts through the ice leaving it intact.

Note that copper wire has been used due to its high thermal conductivity. If a poor thermal conductor like cotton string was used, it would not cut through the ice block.

The process by which water refreezes is referred to as **regelation.**

The effects of high pressure on the melting point are applicable in ice skating and joining two pieces of ice blocks together. The weight of the skater acts on the thin blades of the skates exerting high pressure on the ice. The ice underneath thus melts, forming a thin film of water over which the skater slides.

When two ice cubes are pressed hard against each, the high pressure between them lowers the melting point of the ice at the point of contact. When the pressing force is withdrawn, water recondenses and the two cubes are joined together.

2. The presence of impurities lowers the melting point of a material. This is the reason behind spreading salt on roads and paths during winter in cold regions. This will prevent freezing on the roads.

Boiling

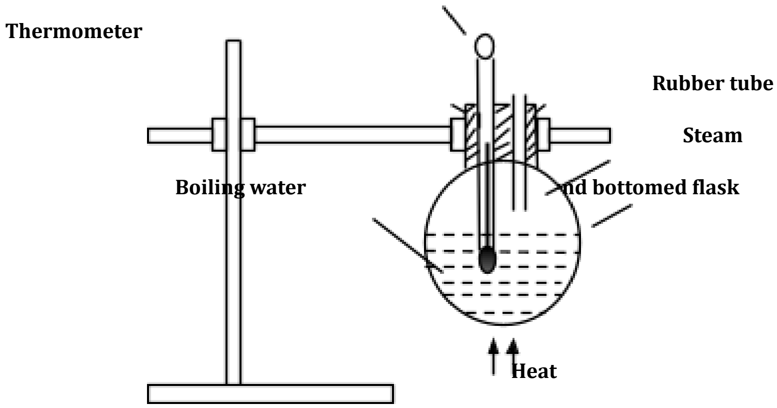
Generally:

-The presence of impurities in a liquid raises the boiling point of the liquid.

-An increase in pressure raises the boiling point of the liquid.

The effects of pressure on boiling point may be illustrated by the set ups below:

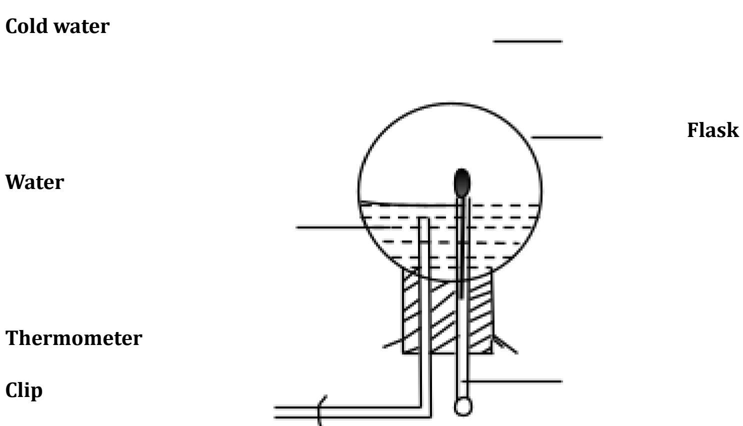
Effects of increased pressure on boiling point



The heating is done until water starts to boil. The temperature at which water boils is noted. When the rubber tube issuing steam is squeezed momentarily, the reading on the thermometer is observed to rise and boiling reduces. Note that closing the tube raises the vapour pressure within the flask. This makes it difficult for the molecules from the surface of the liquid to escape, raising the boiling point of the liquid.

The effect of high pressure on boiling point is applied in a pressure cooker. Here the pressure is raised which raises the boiling point of water hence the food is cooked at a higher temperature.

Effects of reduced pressure on boiling point



Water is first heated to boiling. The flask is then turned upside down and cold water poured over it. It would be observed that when heating stops, boiling also stops. When cold water is poured over the flask, the water inside the flask begins boiling again although its temperature is below the boiling point.

The cold water condenses the steam reducing vapour pressure inside in the flask. Hence a decrease in pressure lowers the boiling point of a liquid.

24.5: Boiling and Evaporation

When a liquid is heated, the molecules close to the surface may gain sufficient kinetic energy to break away from the forces of attraction between the neighboring molecules and escape. This is called **evaporation.** Evaporation takes place at any temperature, even below the boiling point of the liquid.

Factors affecting rate of evaporation

a) Temperature

Increase in temperature of the liquid enhances evaporation. This is why clothes dry faster on a hot day.

b) Surface area

When the surface area is increased, the molecules of the liquid have greater chance of escaping. Hence a wet cloth would dry faster when it is spread out than when it is folded.

c) Humidity

When there is high amount of water vapour in the atmosphere, it becomes difficult for the molecules to escape. This is why clothes take longer to dry on a humid day.

d) Draught/moving wind

Moving air above the surface of the liquid sweeps away the escaping molecules. Thus evaporation is enhanced by the passing air.

DIFFERENCES BETWEEN BOILING AND EVAPORATION

Evaporation	**Boiling**
Occurs at all temperatures	Occurs at a fixed temperature
Occurs at the surface of the liquid	Takes place throughout the liquid
No bubbles are formed	Bubbles are formed in the liquid
Decrease in atmospheric pressure increases the rate of evaporation	Decrease in atmospheric pressure lowers the boiling point of the liquid

Evaporation has a cooling effect which is applied in sweating in human beings and animals, cooling of water in porous pots and the refrigerator.

When water evaporates, it absorbs the latent heat from the body causing a cooling effect. Different animals have different ways by which they cool their bodies. For instance, dogs expose their tongues when it is hot while the muzzle of a cow becomes more wet when it is hot. Both these are to increase the rate of evaporation thereby cooling the body.

A porous pot has tiny holes which allow water to seep out slowly. As the water evaporates, it absorbs the latent heat causing a cooling effect.

TOPIC 25: GAS LAWS

Gas laws looks at the relationship between temperature, volume and pressure of gases.

25.1: Boyle's law

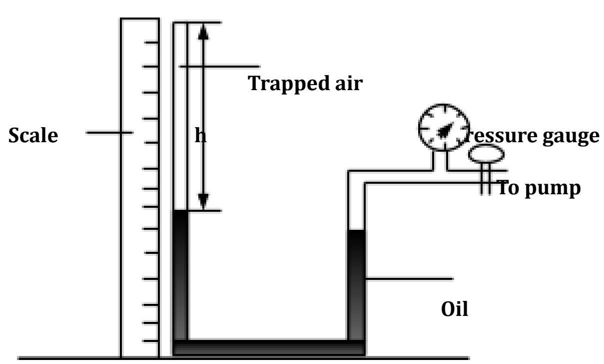
In this law, temperature of the gas is kept constant. Boyle's law states: the pressure of a fixed mass of a gas is inversely proportional to the volume, provided temperature is constant.

$P \alpha 1/V$

$P=k/V$

$PV=$ constant.

The following set up can be used to illustrate Boyle's law:



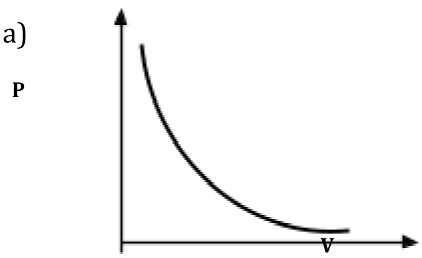
When pressure is exerted on the oil, the trapped gas (usually air) is compressed and the column h reduces. The pressure is measured using the pressure gauge. Since the cross-section area of the glass tube is uniform, the column h can be taken to represent the volume of the trapped gas (air).

Several values of pressure, P and volume, h are collected and recorded.

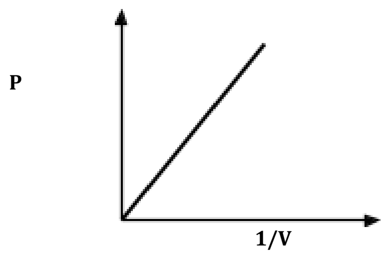
Pressure, P (Pa)	Volume, h(cm)	1/v (or 1/h m^{-1})	PV

A graph of pressure against volume is a curve as shown in (a) below:

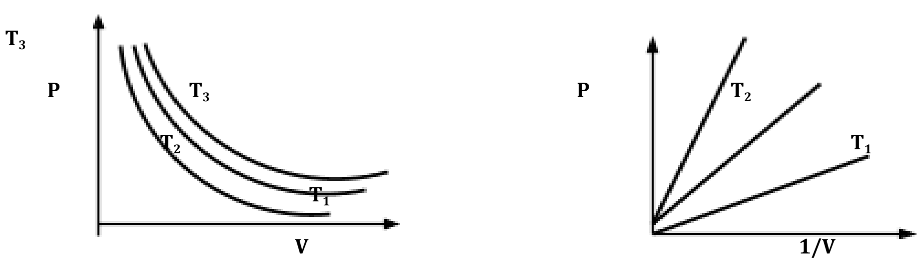
a)



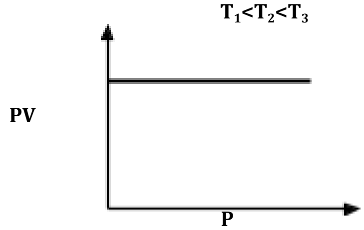
b)



A graph of P against 1/V is a straight line through the origin as shown in (b) above while a graph of PV against P is a straight line parallel to the x-axis. If the experiment is repeated at different temperatures, similar curves to the above will be obtained. This isshownbelow:



$T_1 < T_2 < T_3$



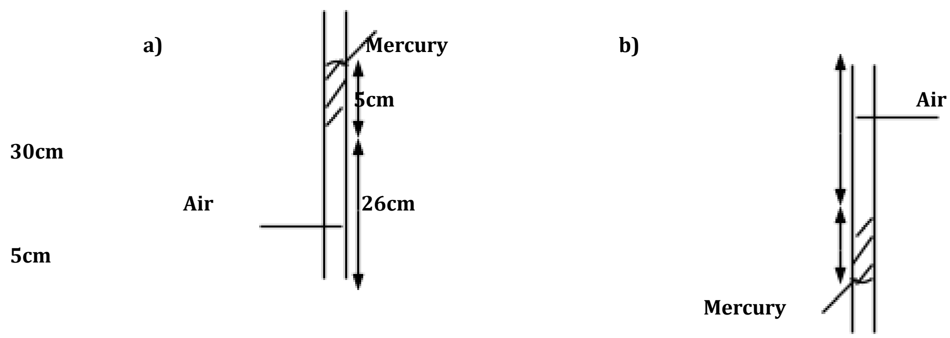
Hence for a given mass of a gas, $P_1V_1 = P_2V_2$

Molecular explanation of Boyle's law

When a gas is put in a closed container, the gas molecules collide with walls of the container generating gas pressure. When the volume of the fixed mass of gas is reduced, the number of collisions per unit time and therefore the rate of change of momentum will increase. Consequently the gas pressure is raised. Hence a reduction in volume leads to an increase in the gas pressure.

Example 25.1

1. A gas occupies a volume of $1.6cm^3$ at a pressure of $1.5 * 10^5Pa$. find the volume it will occupy at a pressure of $3.0 * 10^5Pa$ if the temperature is kept constant.
   $P_1V_1=P_2V_2$
   $V_2= (1.5*10^5*1.6*10^{-6})/ (3.0*10^5) = 8.0*10^{-7}m^3$ or $0.8cm^3$
2. A column of air 26cm long is trapped by mercury thread 5cm long as shown in (a) below. When the tube is inverted as shown in (b), the air column becomes 30cm long. What is the value of the atmospheric pressure?



In (a), the gas pressure = $P_{Atm} + h\rho g$

In (b), the gas pressure = PAtm – $h\rho g$

Let the atmospheric pressure be x metres of mercury.

From Boyle's law, P1V1=P2V2

$(x+0.05)\rho g*0.26=(x-0.05)\rho g*0.3$

$0.26x+0.013=0.3x-0.015$

$0.04x=0.028$

$X=0.028/0.04$

$=0.7m$ (or 70cm)

Hence the atmospheric pressure=70cmHg.

3. The table below shows the results obtained in an experiment to study the variations of the volume of a fixed mass of a gas with pressure at constant temperature:

Pressure, P(cmHg)	**60**	**……**	**90**	**……**
Volume, cm^3	**36**	**80**	**……**	**40**

Fill in the missing values.

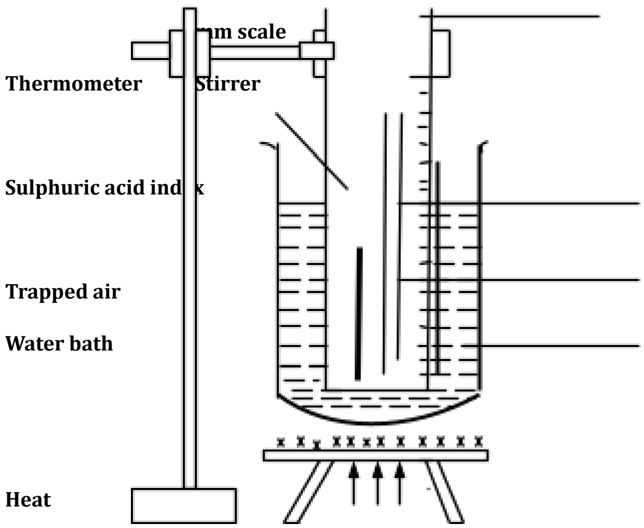
25.2: Charles' law

This law looks at the relationship between temperature and volume of a given mass of gas at constant pressure. It is obvious that when a gas is heated it expands i.e. increases in volume. The law states: the volume of a fixed mass of a gas is directly proportional to its absolute temperature provided the pressure is kept constant.

i.e. $V \alpha T$

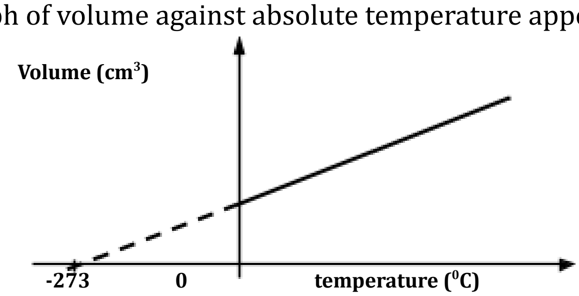
$V=kT$ or V/T = Constant

The set-up below can be used to verify Charles' law:

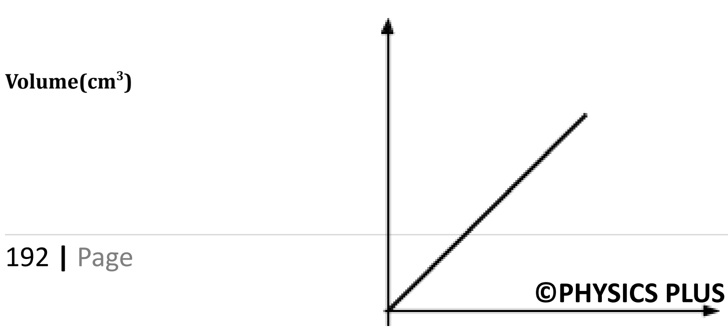


When the gas (trapped air) is heated in a water bath, it increases in volume. This is showed by an increase in the column h of the trapped air. Thus an increase in temperature of the gas causes an increase in its volume.

A graph of volume against absolute temperature appears as shown below:



If the graph is extrapolated, it cuts the x-axis at -273^0C. at this temperature, the gas is assumed to have a volume equals to zero. This is the lowest temperature a gas can ever fall to and is called the **absolute zero**. A temperature scale based on the absolute zero is referred to as the **absolute or Kelvin scale**. On this scale, the temperature must be expressed in Kelvin.



0Absolute temperature (K)

For a given mass of a gas, $V_1/T_1 = V_2/T_2$

This equation ONLY holds when the temperature is expressed in Kelvin.

Molecular explanation of Charles' law

When the temperature of a gas is increased, its molecules gain kinetic energy and move faster. This increases the rate of collision with walls of the container and hence increased pressure. However, since in Charles' law, pressure must be constant, the volume of the container must be increased accordingly so that the gas molecules can cover larger distance before colliding with the walls of the container. This would keep the gas pressure constant although its temperature is raised.

Example 25.2

1. A gas occupies a volume of $125cm^3$ at 15^0C and 755mmHg pressure. Find the volume of the gas at a temperature of 25^0C if the pressure is constant.

   $V_1/T_1 = V_2/T_2$

   $125/(15+273) = V_2/(25+273)$

   $V_2=(125*298)/288 =129.34cm^3$.

2. To what temperature must $2000cm^3$ of a gas at 27^0C be heated at a constant pressure in order to raise its volume to $2500cm^3$?
   $V_1/T_1 = V_2/T_2$

   $T_2=(2500*300)/2000 =375K$ or 102^0C.

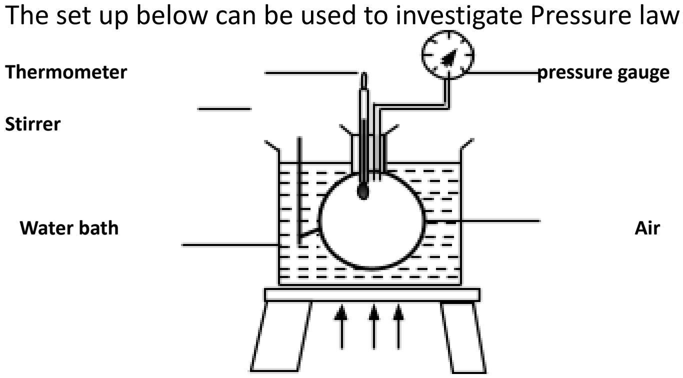
25.3: Pressure law

Raising the temperature of a fixed mass of a gas at a constant volume increases the average kinetic energy of the gas molecules. Pressure law states: ***the pressure of a fixed mass of a gas is directly proportional to its absolute temperature at a constant volume;***

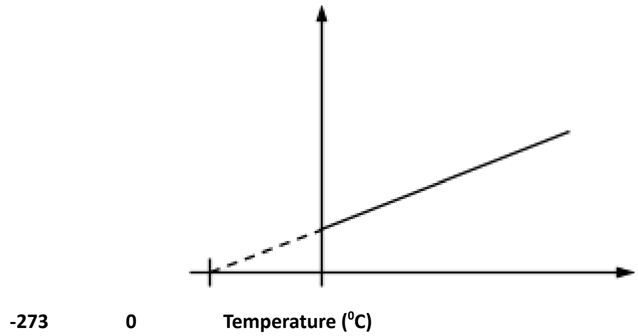
$P \alpha T$ $P=kT$ or $P/T=k$

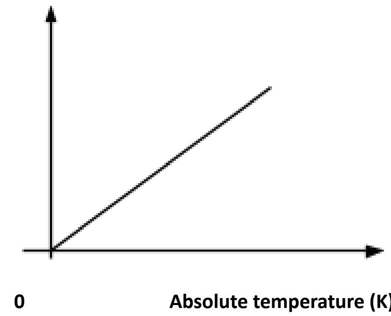
Thus at constant volume, $P_1/T_1= P_2/T_2$

The set up below can be used to investigate Pressure law:



Several values of temperature and the corresponding pressures can be collected and used to plot a graph of pressure against absolute temperature. The graph will appear as shown below:





Example 25.3

1. A tin closed with an airtight lid contains air at a pressure of $1.0*10^5$Pa at a temperature of 12^0C. If the temperature at which the lid opens is 88^0C, determine the pressure attained by the gas.
   $P_1/T_1= P_2/T_2$
   $P_2=[1.0*10^5*361]/285$ =126,666.67Pa

The three laws combined can be expressed as; PV/T =constant, k Or simply

$P_1V_1/T_1= P_2V_2/T_2$

The above equation is referred to as the **equation of state**. In general for a fixed mass of a gas, PV/T=a constant. If 1 mole of the gas is used, then;

PV/T= R, where R is the **universal gas constant**.

Example 25.4

1. A gas occupies a volume of $200cm^3$ at 25^0C and 760mmHg. Find its new volume at -23^0C and 750mmHg.
   $P_1V_1/T_1= P_2V_2/T_2$

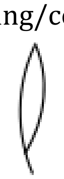
$V_2=[P_1V_1T_2]/P_2T_1=[760*200*250]/[750*298]$ $=170cm^3$

TOPIC 26: THIN LENSES

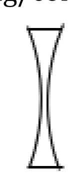
26.1: **Introduction**

A lens is generally a transparent material having at least one curved surface. A lens works by way of refraction of light. There are two common types of lenses:

1. Converging/convex lens- it is thicker at the middle.

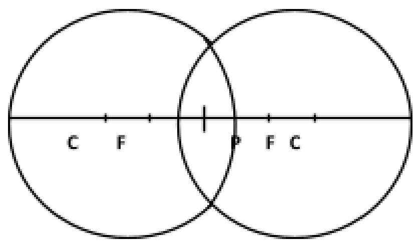


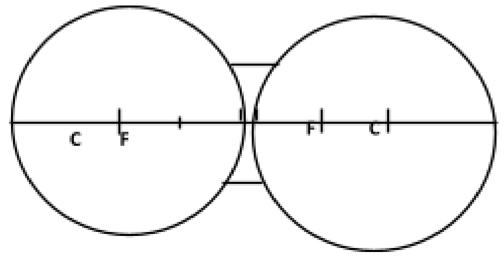
2. Diverging/concave lens- it is thinner at the middle.



26.2: Terms used

a) Centre of curvature, C- it is the centre of the sphere of which the surface of the lens is part. A lens has two centres of curvature.





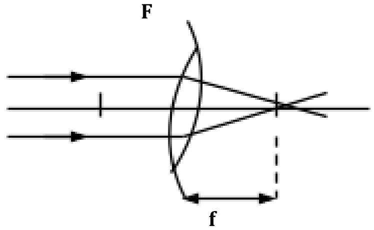
b) Radius of curvature, r- it is the radius of the sphere of which the lens is part.

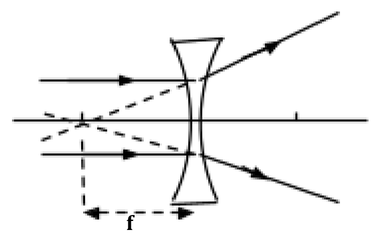
c) Principal axis- a straight line joining the two centres of curvature.

d) Optical centre, P- it is a point on the principal axis midway between the lens surfaces.

e) Principal focus, F- for a converging lens, it is the point along the principal axis at which rays parallel and close to the principal axis converge after refraction by the lens. For a diverging lens, it is the point along the principal axis from which rays parallel and close to the principal axis seem to diverge from after refraction by the lens.

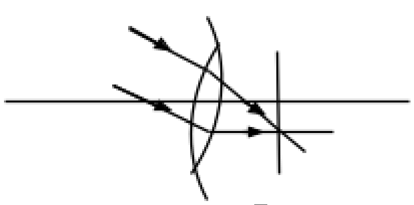
F F

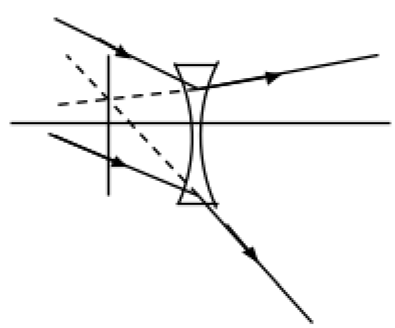




f) Focal length, f- it is the distance between the optical centre and the principal focus. It is real for a converging lens and virtual (negative) for a diverging lens.

g) Focal plane- when parallel rays which are not parallel to the principal axis are incident on a lens, the rays converge at or appear to diverge from a point which is perpendicular to the principal axis and passes through the principal focus, F. this plane is called the focal plane.

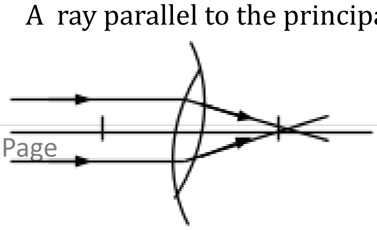


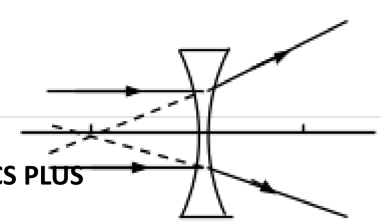


26.3: Lens ray diagrams

There are three main rays in ray diagrams as shown below:

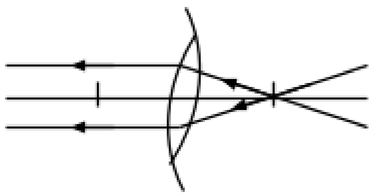
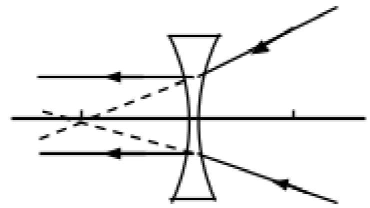
a. A ray parallel to the principal axis





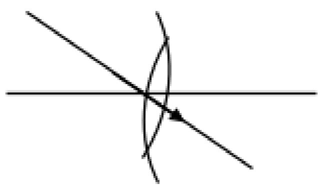
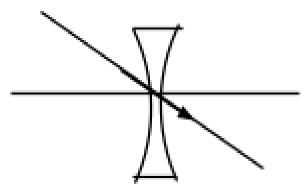
For a converging lens, the rays converge at F after refraction while for a diverging lens, the rays appear to diverge from F after refraction.

b. A ray passing through or appearing to pass through F.

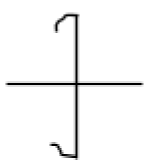
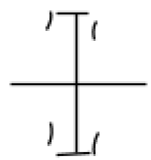
In both cases, the refracted rays are parallel to the principal axis.

c. A ray directed to the optical centre of the lens.

In both cases, the rays pass undeviated.

NB: in ray diagrams, the following symbols are used for the two lenses:

Converging lens **Diverging lens**

Note that in ray diagrams:

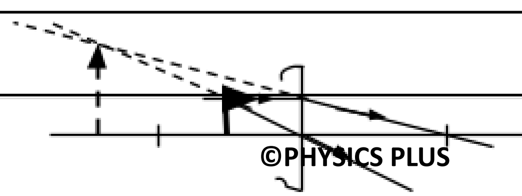
1. Real rays, real objects and real images are represented using continuous lines.
2. Virtual rays and virtual images are represented using broken lines.
3. To locate an image, there must be at least two rays intersecting, whether real or virtual.

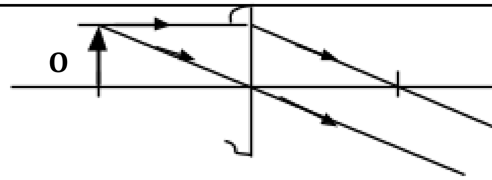
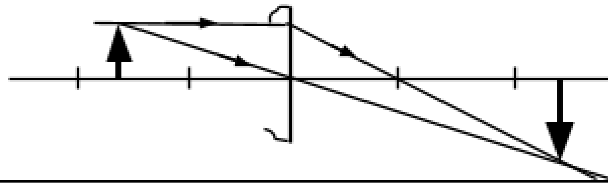
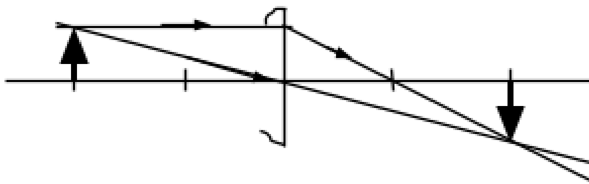
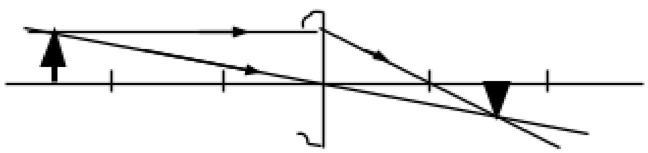
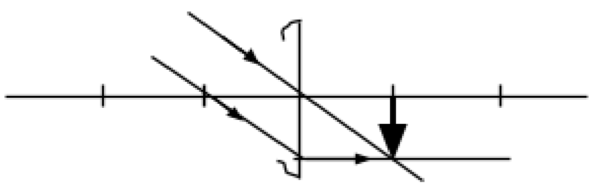
Sometimes, a scale may be used in ray diagrams. If used, then the scale chosen for object and image distances need not be necessarily equal to that of the object and image heights but the two must be given on the diagram.

26.4:Image formation by thin lenses

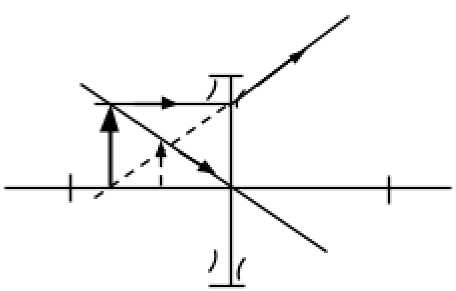
26.4.1:Image formation by a converging lens.

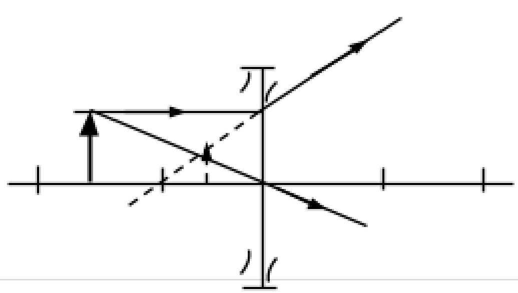
This is summarized by the table below:

Position of object	Ray diagram	Characteristics of image
Between F and the lens		Image is: - Virtual

	I O	- Upright/erect - Magnified - On same side as the object
At F	 	Image is: - Real - Inverted - At infinity
Between F and C	O I	Image is: - Real - Inverted - Magnified - Beyond C
At C		Image is: - Real - Inverted - Same size as object - At C
Beyond C		Image is: - Real - Inverted - Diminished
At infinity		Image is: - Real - Inverted - Diminished - At F

26.4.2:Image formation by a diverging lens





Generally, a diverging lens forms a virtual, upright and diminished image regardless of the position of the object.

26.5:The lens formula and magnification

The equation 1/f=1/u + 1/v where f is the focal length of the lens, u is the object distance and v the image distance, is called the lens formula. The equation takes into account the signs of u, v and f and holds for both the converging and diverging lens.

The ratio of the image size to the object size is called magnification of the lens. When the magnification is less than one the image is diminished while when it is more than one, the image is magnified. When the magnification is one, then the object and image are of the same size.

Magnification can also be obtained from the ratio of image distance to object distance i.e.

Magnification, m= image height/object height = image distance, v/object distance, u

Example 26.1

1. An object is placed in front of a converging lens of focal length 12cm.determine the position of the image and the magnification of the image. Hence state the nature of the image formed when the object distance is:

a. 16cm (**ans: v=48cm,image real and magnified)**

b. 8cm (**ans: v= -24cm, image is virtual and upright)**

2. An object is placed 4cm in front of a diverging lens of focal length 6cm. find the position and magnification of the image formed. Hence state its nature.

3. An object 10cm high is placed 30cm in front of a converging lens of focal length 20cm. determine the position and height of the image by:

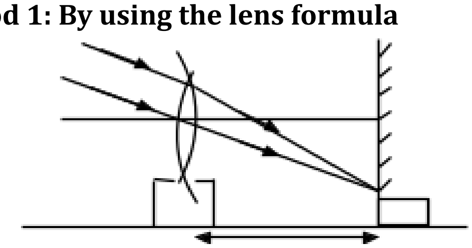
a. Calculation

b. Scale drawing

26.6:Determination of the focal length of a converging lens.

Method 1: By using the lens formula

Screen

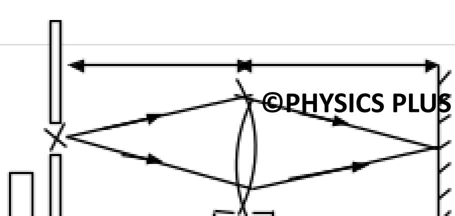


f

- Focus clearly the image of a distant object like tree or window on the screen by adjusting the position of the lens appropriately. Measure the distance between the lens and screen,

  f= ………………..cm.

- Now set up the apparatus as shown below:



u v

- Set the object distance, u=65cm. adjust the position of the screen until a sharp image of the object is observed. Measure the distance, v.
- Reduce the object distance in steps of 10cm and measure the corresponding values of v.
- Record your results and complete the table the table below:

u(cm)	65.0	55.0	45.0	35.0	25.0
V(cm)					
1/u(cm^{-1})					
1/v(cm^{-1})					

- Plot a graph of 1/u against 1/v. hence determine the focal length, f of the lens.

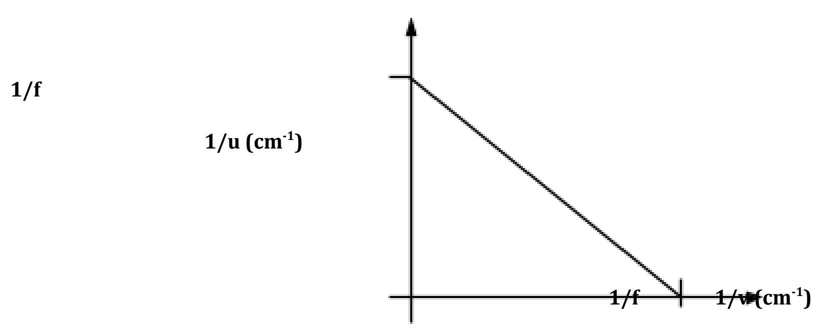
Note that from the lens formula, 1/f=1/u + 1/v

Making 1/u the subject of the formula, we have 1/u=-1/v +1/f.

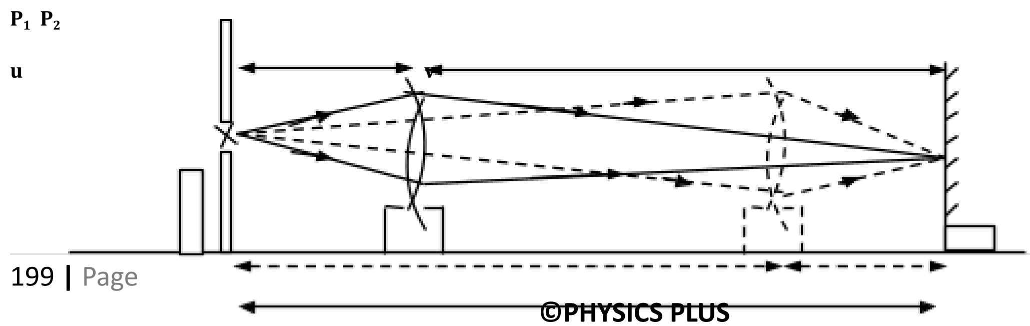
When 1/v=0, 1/u=1/f i.e. y-intercept

And when 1/u=0, 1/v=1/f i.e. x-intercept.

Therefore the graph is a straight line whose slope is -1 and the y-intercept and x-intercept give 1/f. if the values of the two intercepts are different, then their average is obtained.



Method 2: Displacement method



u_1 v_1

S

- Estimate the focal length of the lens by focusing a distant object.

  f=………..cm

- Set up the apparatus as shown above. The distance, **S** should be equal to or more than 4f.adjust the position of the lens to a position P_1 where a sharp image of the object is obtained on the screen. Measure and record u and v.
- Keeping **S** constant, move the lens to another position P_2 where another clear but diminished image of the object is formed on the screen. Measure and record u_1 and v_1.
- Increase the value of **S** in steps of 5cm and repeat steps 2 and 3 above. Complete the table below:

S(cm)					
U(cm)					
V(cm)					
U_1					
V_2					
d=v-u					
(S^2-d^2)(cm^2)					

- Plot a graph of (S^2-d^2) against S.
- Given that the equation S^2= 4fS +d^2 satisfies the graph drawn, determine the value of f.

Note that S= u + v ………………………………………….(i)

And d=u_1-u or d=v - v_1

But u_1=v and v_1=u

Therefore, d= v – u ……………………………………………. (ii)

Adding equations (i) and (ii), we obtain;

S + d = u + v + v – u = 2v

V= (S + d)/2 ………………………………………………… (iii)

Also subtracting equation (ii) from (i), we get;

S – d = u + v – v + u = 2u

u= (S – d)/2 ………………………………………………….. (iv)

Substituting equations (iii) and (iv) in the lens formula, we get;

$1/f= 1/\{(S-d)/2\} + 1/\{(S+d)/2\}$

$1/f=2/(S-d) + 2/(S+d)$

$1/f=4S/(S^2-d^2)$

$(S^2-d^2) = 4Sf$

Hence a graph of (S^2-d^2) against S is a straight line through the origin and whose slope equal to 4f.

The two positions P_1 and P_2 are known as conjugate points.

26.7: Other possible graphs from the lens formula

1. From the lens formula; $1/f=1/u +1/v$

   $1/f= (v+u)/uv$

   $uv= (u+v)f$

   Hence a graph of uv against (u+v) is a straight line through the origin and whose slope equal to the focal length, f of the lens.

2. Also, from the lens formula; $1/f= 1/u + 1/v$

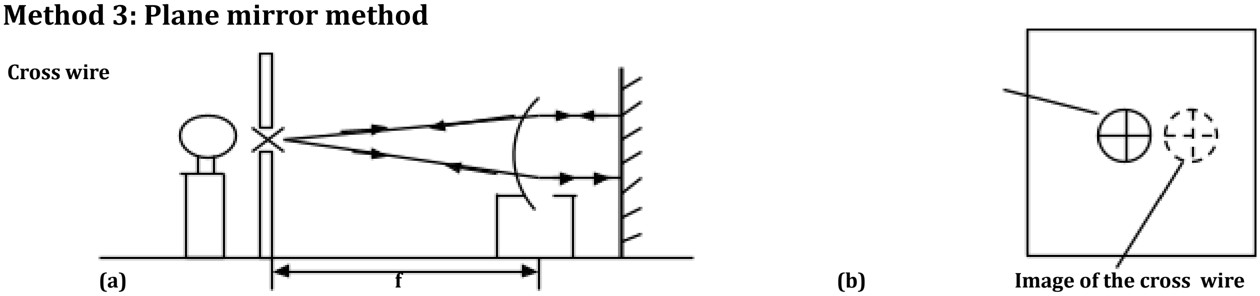
   Multiplying through by v, we obtain $v/f= v/u + 1$

   But $v/u= m$

   Therefore, $m=v/f – 1$

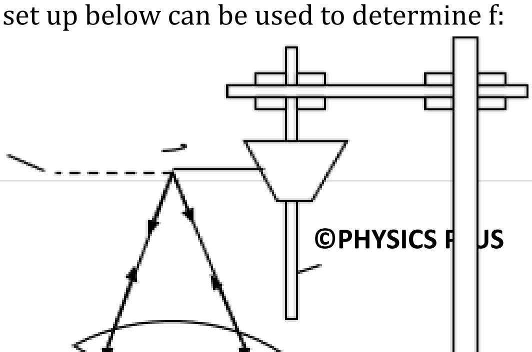
   Hence a graph of m against v is a straight line whose slope equal to 1/f and the y-intercept= -1.

Method 3: Plane mirror method



Adjust the position of the object until a sharp image of the cross wire is formed alongside (close to) the object cross wire as shown in (b) above. The distance between the object and the centre of the lens gives the focal length f of the lens.

Alternatively, the set up below can be used to determine f:



Pin **Sliding cork**

Glass rod

Lens

Plane mirror

Adjust the position of the cork up and down until the pin and its image just coincide i.e. no parallax between the object pin and its image. The distance between the centre of the lens and the pin gives the focal length of the lens.

26.8: Power of a lens

It is the measure of the refracting ability of the lens. It is expressed as the reciprocal of the focal length

i.e. power of a lens= 1/focal length

It is measured in dioptre (D).

The shorter the focal length the higher its refracting ability. The power of a converging lens is positive while that of a diverging lens is negative.

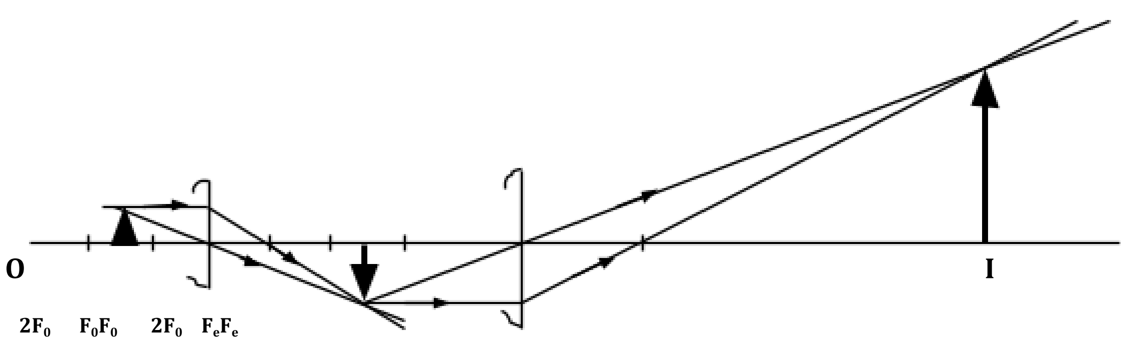
26.9:Applications of lenses

- **A simple microscope**

It is also known as a magnifying glass. When a converging lens is used such that the object is found between its principal focus and the lens, it forms a virtual, upright and magnified image. When used this way it serves as a simple microscope.

- **A compound microscope**

It consists of two converging lenses, objective lens and eyepiece lens both of short focal lengths. The lens closer to the object is called the objective lens while that closer to the eye is called the eyepiece lens. The focal length of the eyepiece lens is longer than that of the objective lens. The object is found between F and C of the objective lens.



The first image formed by the objective lens is real, inverted and magnified. This image then acts as the object for the eyepiece lens. The eyepiece lens forms a final image which is greatly magnified. Assuming the magnification of the objective lens is m_o and that of the eyepiece lens is m_e , then the total magnification of the compound microscope $m= m_o * m_e$.

Example 26.2

1. In a compound microscope, the focal length of the objective lens is 2.0cm and that of the eyepiece lens is 2.2cm and they are placed at a distance of 8.0cm. a real object of size 1.00mm is placed 3.0cm from the objective lens.

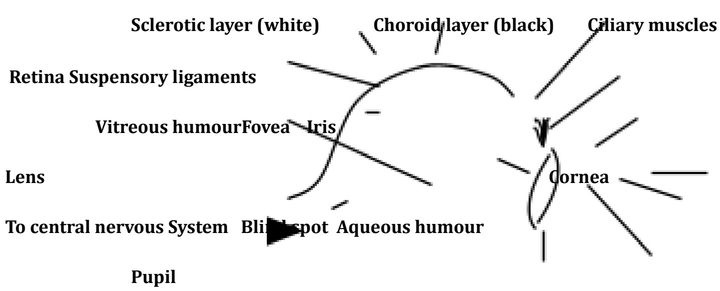
a) Use the lens formula in turn for each lens to find the position of the final image formed.

b) Calculate the magnification produced by the arrangement of these lenses and the size of the final image viewed by the eye.

- **The lens camera**

A camera is a device that is used to take photographs. It consists of a converging lens, a light- sensitive film enclosed in a light-tight box blackened on the inside and a shutter. Light from the object enters the camera through the shutter which closes automatically after a given length of time. The amount of light reaching the lens is controlled by the diaphragm (stop). The light reaching the lens is refracted to form a real, inverted and diminished image on the film. To clearly focus the image, the distance between the lens and film is adjusted accordingly. The film has some light- sensitive chemicals which change on exposure to light. This can then be developed and printed to get a photograph.

- **The human eye**



The human eye is a natural optical instrument. It comprises of the following parts:

- Sclerotic layer- encloses the eye. The front part (cornea) is transparent to allow in light. It is the white part of the eye.
- Aqueous humour- it is a clear fluid/liquid found between the cornea and the lens. This helps to maintain the shape of the eye.
- Iris- it is responsible for the colour of the eye. It has the pupil in the middle which allows for passage of light. By changing the size of the pupil, the iris controls the amount of light entering the eye.
- Lens- it is a natural converging lens. With the help of the ciliary muscles, its focal length can be adjusted for fine focusing.
- Vitreous humour- it is a jelly-like substance and transparent in nature found between the lens and retina.

- Retina- images are formed here. It has light-sensitive cells.
- Fovea- it is the central part of the retina. This is where the eye has the best details and colour vision.
- Blind spot- has cells which are non-sensitive to light.
- Ciliary muscles- they suspend/support the lens. It is also responsible for controlling the shape of the lens.

When the muscles relax, the focal length of the lens increases. This enables the eye to focus a distant object. Contraction of the muscles on the other hand reduces tension in the lens, thus reducing its focal length. This enables it focus near objects.

This automatic adjustment of the eye lens to bring to focus on the retina images of both distant and near objects is referred to as **accommodation.**

The closest/shortest distance a normal eye can focus clearly is known as its **near point** while the farthest distance a normal eye can focus clearly is known as its **far point**. For a normal eye, the near point is usually 25cm.

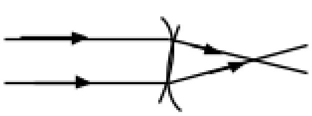
Note that the distance between the retina and the eye lens is always constant.

26.9.1:Eye defects

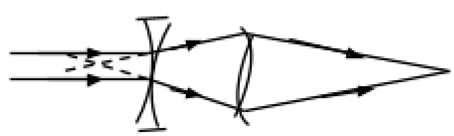
Despite the adjustments made by the eye, some eyes cannot produce clear images within the normal range of vision. There are two common eye defects namely myopia (shortsightedness) and hypermetropia (long-sightedness).

Myopia (shortsightedness)

Having this defect means clear vision for near objects but images of distant objects are formed in front of the retina. The cause of the defect is the eyeball being too long or shorter focal length. The defect is corrected by using a diverging lens of appropriate focal length so that the rays reaching the eye lens appear as if they are coming from a near object.



Defect

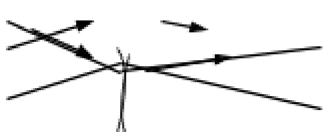


Correction

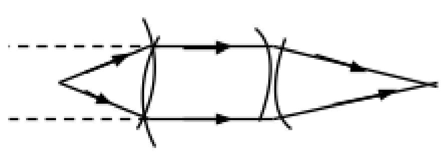
Hypermetropia (long-sightedness)

A person who is long-sighted has clear vision of distant objects but cannot see clearly closer objects clearly. It is caused by the eyeball being too short or longer focal length so that the image of a closer object is formed behind the retina. The defect is corrected by using a converging lens of appropriate focal length.

Defect



Correction



26.9.2: Similaritiesanddifferences between the eye and lens camera

Similarities

1. Both use converging lenses.
2. In both cases, the amount of light allowed in can be controlled. The eye does it through the iris while the camera does this through the diaphragm.
3. In both, a real, inverted and diminished image is formed. For the eye, the image is formed on the retina while for the camera, it is formed on a light-sensitive film.
4. In both cases the inner part is black; for the eye, there is the choroid layer which is black and for the camera, the inner part is painted black. This is to absorb stray rays.

Differences

1. The focal length of the eye lens changes while that of the lens camera is constant.
2. The distance between the lens and film in a lens camera can be varied by zooming while the distance between the eye lens and retina is constant.
3. A camera can take only one photo at a time when the shutter is open while the eye forms constantly changing pictures.

TOPIC 27: UNIFORM CIRCULAR MOTION

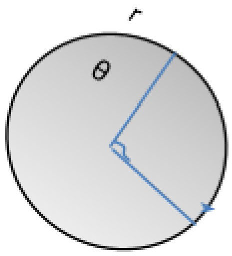
27.1: Introduction

This is motion around a curved path in which the speed of the particle is uniform. Although the speed is the same, velocity of the body keeps varying since there is a constant change in direction. The velocity of such a body at any instant is always directed along the tangent of the curve such that if the body breaks away, it moves along the tangent.

Since the velocity is constantly changing, a body undergoing uniform circular motion is always accelerating.

27.2: Terms used in uniform circular motion

- Angular displacement, ϴ- is the angle swept at the centre by an arc. It is measured in radian, Θ^c .



ϴ=(arc length, S/radius, r) radians

When S= r, then ϴ= S/S =1 radian.

A radian is therefore the angle subtended at the centre of a circle by an arc length equal to the radius of the circle.

Note that $2\pi^c = 360^0$

- Angular velocity, ω- it is the rate of change of angular displacement.

$\omega = \Delta\Theta/\Delta t$

The SI unit of angular velocity is radian per second ($rads^{-1}$).

Remember ϴ= S/r

Therefore, $\Delta\Theta = \Delta S/r$, since r is constant.

Thus $\omega = \Delta\Theta/\Delta t = \Delta S/r\,\Delta t$

But $\Delta S/\Delta t$ = velocity, v

Hence ω= v/r

Thus a body in circular motion has both linear and angular velocities.

- Period, T- it is the time taken by a body to make one complete revolution.

Period, T= circumference (2πr)/linear velocity, v.

Therefore T= 2πr/rω since v= rω

And T= 2π/ω

Recall frequency, f= 1/T

Therefore f= 1/(2π/ω) =ω/2π

Or simply $\omega = 2\pi f$

- Angular acceleration, α- it is the rate of change of angular velocity.

$\alpha = \omega/t$

- Centripetal acceleration, a- it is the acceleration of a body undergoing uniform circular motion and is always directed towards the centre of the circle.

It is given as $a = v^2/r$.

But $v = r\omega$

Therefore $a = (r\omega)^2/r = r\,\omega^2$.

- Centripetal force, F_c- it is the force which keeps the body on its circular path. It is expressed as $F_c = m\,v^2/r = ma$

Where m- is the mass of the particle
v- is its linear velocity
r- is the radius of the path
a-is the centripetal acceleration

Also, $a = r\,\omega^2$. Thus $F_c = m\,r\,\omega^2$

Hence, there are three factors affecting the centripetal force. These are:

- Mass of the body
- Radius of the path
- Angular velocity of the body.

Example 27.1

1. A particle moves at an angular velocity of 10π rads^{-1} along a circular path of radius 20cm. calculate its linear speed.

$v = r\omega = 0.2*10\pi = 6.283$m/s.

2. A car of mass 100kg moves round a circular track of radius 100m with a linear speed of 20m/s. calculate:

a) The angular velocity

$\omega = v/r = 20/100 = 0.2$rads^{-1}

b) The centripetal force

$F_c = m\,v^2/r = (100*20^2)/100 = 400$N

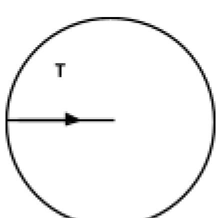
c) The centripetal acceleration

$a = v^2/r = 20^2/100 = 4$m/s^2

3. Calculate the period and frequency of a particle in a circular path moving at an angular velocity of 4π rads^{-1}.

$T = 2\pi/\omega = 2\pi/4\pi = 0.5$s
$f = 1/T = 1/0.5 = 2$Hz

27.3: Motion in a horizontal circle



In a horizontal circle, the tension in the string provides the centripetal force and is directed towards the centre of the circle.

$F_C = T = m\,v^2/r$

For a fixed radius, the tension is constant all round.

Example 27.2

1. A stone of mass 2kg is whirled in a horizontal circle of radius 0.5m. The tension in the string was found to be 64N. Determine the velocity of the stone in the circle.

$V=(r F_C/m)^{1/2} =(0.5*64/2)^{1/2}= 4m/s$

2. A stone of mass 0.6kg attached to a string of length 0.5m is whirled in a horizontal circle at a constant speed. If the maximum tension in the string is 30N before it breaks, calculate:

a) The maximum speed of the stone

$V=(r F_C/m)^{1/2} =(0.5*30/0.6)^{1/2} = 5m/s.$

b) The maximum number of revolutions per second it can make

$f=\omega/2\pi =v/2\pi r$

$= 5/2\pi*0.5 =1.593Hz.$

Assignment 27.1

1. A stone of mass 5kg is whirled at the end of a string 5m long in a horizontal circle. If it is whirled at 5revolutions per second, calculate:

   a) The angular velocity

   b) The linear velocity

   c) The centripetal acceleration

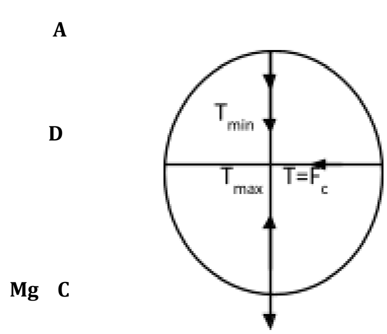
   d) The tension in the string

2. An object of mass 4kg moves round a circle of radius 6m with a constant speed of 12m/s. calculate:

a) Its angular velocity

b) The centripetal force

27.4: Motion in a vertical circle



When a body is whirled in a vertical circle, two forces act on it. These are the weight, mg of the body and the tension, T on the string. The tension varies depending on the position of the body. It is always directed towards the centre of the circle.

At the top of the circle, both the weight of the body and the tension are directed to the centre. Hence the centripetal force equal to the sum of the weight of the body and the tension.

i.e. $F_c=T_{min}+ mg$

Or $T_{min}=F_c-mg= mv^2/r -mg$

At the bottom of the circle, the centripetal force is given by the difference between the weight and the tension; $F_c=T_{max}-mg$

Or $T_{max}=F_c+mg = mv^2/r +mg$

At B and D, the tension in the string provides the centripetal force;

$T=F_c=mv^2/r$

Hence tension is maximum at the bottom and least at the top.

When the tension is minimum, a certain minimum speed must be maintained in order to keep the string taut. Taking minimum tension to be zero, we have;

$0=mv^2/r –mg$ (from the equation of minimum tension)

$mv^2/r =mg$

$V^2=rmg/m = rg$

Therefore, $v=(rg)^{1/2}$

This is the expression for minimum speed that the body must maintain. It is also called the critical speed.

Example 27.3

1. An object of mass 8kg is whirled round rapidly in a vertical circle of radius 2m with a constant speed of 6m/s. Calculate the maximum and minimum tension in the string. Take g=10N/kg.

   $T_{max}=mv^2/r +mg =8(6^2/2 +10) = 224N$

   $T_{min}= mv^2/r –mg =8(6^2/2 -10) = 64N$

2. An object of mass 0.4kg is rotated by a string at a constant speed, v in a vertical circle of radius 1m. If the minimum tension in the string is 3N, calculate :

a) The speed, v

   $T_{min}= mv^2/r –mg$

   $3=0.4(v^2/1 -10)$

   $V^2=(3+4)/0.4 =70/4$

   $V=(70/4)^{1/2}=4.183m/s$

b) Maximum tension in the string

   $T_{max}=m(v^2/r +g) =0.4(70/4 +10)$

= 11N

c) The tension when the string is just horizontal

$T=mv^2/r = 0.4*70/4 =7N$

3. A bob of mass 1kg is moving in a uniform circular path in a vertical plane of radius 1m. If it is whirled at a frequency of 2cycles per second, calculate:

a) The tension in the string when the bob is at the top part of the circle

$T=mv^2/r - mg= mr\omega^2-mg=mr(2\pi/f)^2-mg$

$= (1*1*4*\pi^2/2^2)-(1*10)$

= 147.95N

b) Tension when it is at the bottom

$T= mr(2\pi/f)^2+mg$

$= (1*1*4*\pi^2/2^2) + (1*10)$

= 167.95N

c) At what position is the string likely to break? Why?

At the bottom of the circle. The string experiences maximum tension at this point.

d) What is the minimum speed required to maintain the string under tension.

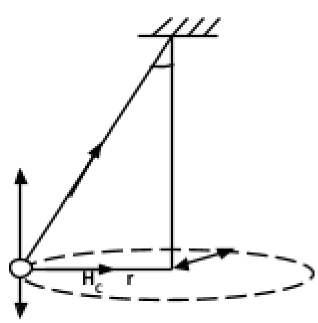
$V=(rg)^{1/2}=(1*10)^{1/2}= 3.162m/s$

27.5: A conical pendulum

Θ

T

V_C



Ball

Mg

When the angular velocity, ω of the ball increases the angle Θ also increases and the ball rises. The horizontal component, H_C of the tension provides the centripetal force;

i.e. $F_C=H_C= mv^2/r$

But H_c=TsinΘ

Therefore TsinΘ= mv^2/r…………………..(i)

Also TcosΘ= mg……………………………….(ii)

Dividing equation (i) by (ii), we get;

TsinΘ/TcosΘ = (mv^2/r)/mg

Therefore tanΘ=v^2 /rg

Also tanΘ=$r\omega^2$/g   (since v=rω)

Example 27.4

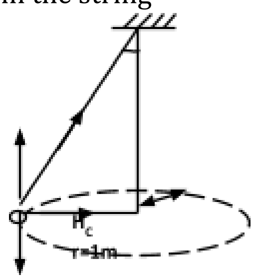
1. A stone of mass 2kg is attached to a string 3m long made to revolve in a horizontal circle of radius 1m. find:

a) The tension in the string

θ

3m

TcosΘ



Mg

Θ=Sin^{-1}(1/3)= 19.47^0

Therefore T=mg/cosΘ =20/cos19.47^0

=21.21N

b) The linear velocity, v

TsinΘ= mv^2/r

Therefore v= $(21.21\sin 19.47^0 * 1/2)^{1/2}$ = 1.88m/s

c) The angular velocity, ω

ω= v/r =1.88/1 =1.88rads^{-1}

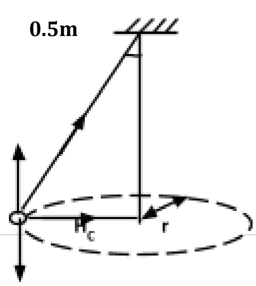
2. A small pendulum bob of mass 250g is suspended by an inelastic string of length 0.5m. The bob is made to rotate in a horizontal circle of radius 0.4m and whose centre is vertically below the point of suspension. Calculate:

a) The magnitude of the component forces acting on the bob due to the tension in the string

0.5m

0.3m

V_c



Mg

V_C=TcosΘ=mg = 0.25*10= 2.5N

But cosΘ=3/5

So T=2.5*5/3 =4.17N

H_C=TSinΘ= (12.5/3)*4/5 = 3.33N

b) The period of rotation of the bob

$TSin\Theta = mr\omega^2$

$\omega = (TSin\Theta/mr)^{1/2} = (3.33/0.4*0.25)^{1/2}$

$= 5.77 rads^{-1}$

But period, T=2π/ω=1.0889s

Assignment 27.2

1. An object of mass 10kg is whirled round a horizontal circle of radius 4m by a revolving string inclined to the vertical. If the uniform speed of the object is 5m/s, calculate:
   - **a)** The tension in the string
   - **b)** The angle of inclination of the string to the vertical

   (**Ans. T=118N, Θ=32^0**)

27.6: Case examples of circular motion

27.6.1: movement of cars round a flat (level) bend

In this case, the centripetal force is provided by the frictional force between the tyres and the road, i.e. $F_R = F_C = mv^2/r$

If the road is slippery, then the frictional force may not be sufficient to provide the required centripetal force. Hence skidding may occur.

To avoid skidding, the speed of the car should not exceed a certain speed limit (critical speed). The critical speed depends on the radius of the bend; the larger the radius the higher the critical speed.

Other factors which affect friction also significant here. These include the nature of the road surface and the nature of the tyres.

Example 27.5

1. A car of mass 1200kg is moving with a velocity 25m/s round a flat bend of radius 150m. Determine the minimum frictional force between the tyres and the road that will prevent the car from sliding off.

$F_R = mv^2/r$ = 1200*25^2/150 = 5000N

2. A glass block of mass 100g is placed in turn at various distances from the centre of a table which is rotating at a constant angular velocity. It is found that at a distance 8cm from the centre, the block just starts to slide off the table. If the frictional force between the block and the table is 0.4N, determine:

a) The angular velocity of the table

$F_R = mr\omega^2$

$\omega = (0.4/0.1*0.08)^{1/2} = 7.07 rads^{-1}$

b) The force required to hold the block at a distance of 12cm from the centre of the table.

$F = mr\omega^2 = 0.1*0.12*50 = 0.6N$

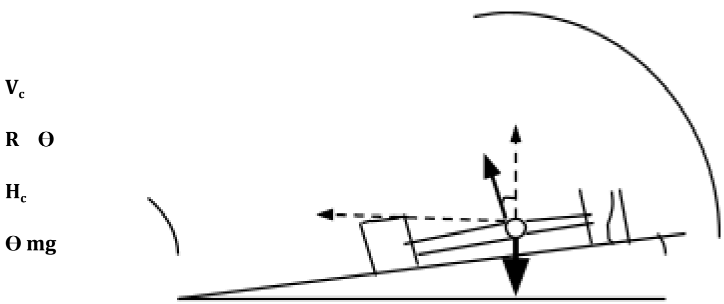
c) A block of mass 200g is now placed 8cm from the centre of the table and the table rotated at the same angular velocity. State with a reason whether or not the block will slide off.

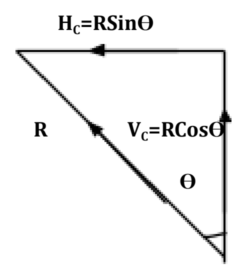
$F_c = mr\omega^2 = 0.2*0.08*50 = 0.8N$

Since the required centripetal force is greater than the frictional force, the block will slide off.

27.6.2: Banked road

Banking refers to raising the outer edge of the road away from the centre of the curve. The angle through which the road is raised is called the angle of banking. Banking minimizes chances of skidding/overturning.





In this case, the centripetal force is provided by the horizontal component of the normal reaction.

Therefore $F_C = RSin\Theta = mv^2/r$………………………… (i)

Also $RCos\Theta = mg$………………………………………….. (ii)

Dividing equation (i) by (ii), we get;

$\tan\Theta = v^2/rg$

Note that for no skidding to occur, $\tan\Theta$ must be greater than v^2/rg ($\tan\Theta > v^2/rg$) and the critical speed, v is given by;

$v = (rg\tan\Theta)^{1/2}$

Hence the critical speed, v depends on the radius, r of the curve and the angle of banking, Θ.

Example 27.6

1. The maximum speed at which a motorist can negotiate a corner on a level ground is 20m/s. if the radius of the curve is 100m, calculate:

a) The coefficient of friction, μ

$F_R = F_C = mv^2/r = \mu R = \mu mg$

Therefore $v^2/r = \mu g$

$\mu = 20^2/100*10 = 0.4$

b) The angle at which the road should be banked to enable the motorist negotiate the corner at a critical speed of 20m/s.

$\tan\Theta = v^2/rg = 0.4$

$\Theta = \tan^{-1}0.4 = 21.8^0$

2. A vehicle moves round a banked balanced road at a speed of 20m/s. if the radius of the road is 50m, calculate the angle of banking.

$\tan\Theta = v^2/rg = 20^2/50*10 = 0.8$

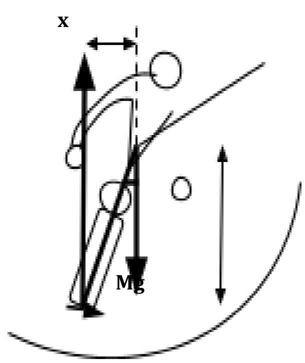
$\Theta = \tan^{-1}0.8 = 38.7^0$

27.6.3: Aircraft banking

An aircraft turning in air achieves banking by lowering one of its wings and raising the other. This enables it turn at extremely high speed without overturning.

27.6.4: A cyclist moving round a circular track.

c.o.g



Θy

F_R

For a flat road surface, frictional force between the tyres and the road provides the centripetal force;

$F_R = F_C = mv^2/r$

In cases when the frictional force is not sufficient, the cyclist is likely to skid. To avoid this, the cyclist is advised to lean inwards. Thus the frictional force, F_R and the normal reaction, R produce a turning effect about the centre of gravity.

Thus taking moments about the c.o.g., we get;

Clockwise moment=Rx

And anticlockwise moment= $F_R y$

For no skidding to occur, clockwise and anticlockwise moments should be equal;

$Rx = F_R y$

$F_R/R = x/y$

But $\tan\Theta = x/y$ and $R = mg$

Therefore $\tan\Theta = F_R/mg = \mu$

Where μ is the coefficient of friction.

Hence skidding only occurs when $\tan\Theta$ is greater than μ ($\tan\Theta > \mu$).

Example 27.7

1. A cyclist who is travelling at 20m/s negotiates a bend of radius 45m. he inclines at an angle Θ to the vertical. Calculate:

a) The centripetal acceleration

$a = v^2/r = 20^2/45 = 8.889 m/s^2$

b) The angle of inclination Θ

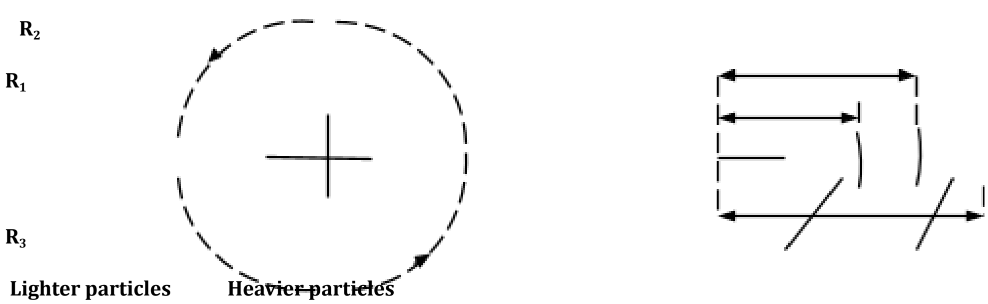
$\tan\Theta = v^2/rg$

$\Theta = \tan^{-1}(20^2/45*10) = 41.63^0$

2.7: Applications of circular motion

27.7.1: Centrifuges

A centrifuge is a device that is used to separate substances of different densities e.g. immiscible liquids or solids suspended in liquids. The mixture is put in tubes which are then set into rotation. At a particular speed, the more dense particles or substance move further away from the centre of rotation while the less dense particles move inwards towards the centre of rotation.



Since centripetal force varies directly as the mass and inversely as the radius, for a larger radius the mass must be higher for the same amount of centripetal force. Hence denser particles are far away from the centre of rotation.

When the rotation stops, the tubes return to the vertical position with the denser particles at the bottom.

27.7.2: Satellites

When two bodies of mass m_1 and m_2 are separated by a distance r, there existsattractional force between them given by; $F = Gm_1m_2/r^2$

Where G is the universal gravitational constant. The above equation is referred to as Newton's law of universal gravitation. Consider a satellite of mass m orbiting the earth at a distance of r metres away. Suppose the mass of the earth is M, then the centripetal force that keeps the satellite on the circular path is provided by the gravitational force of attraction.

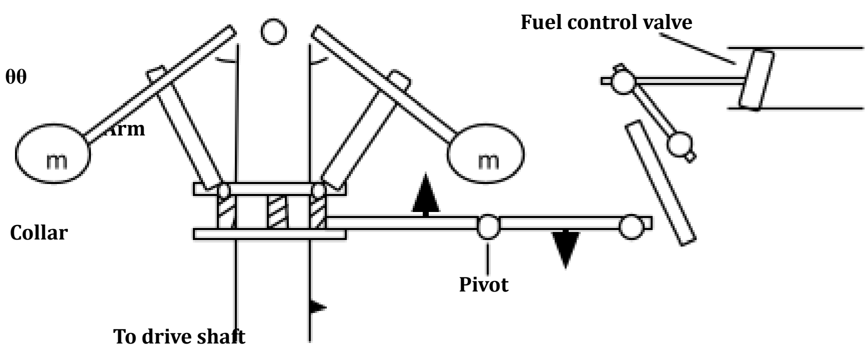
Hence $F_C = mv^2/r = GMm/r^2$

And $v=(GM/r)^{1/2}$

Where v is the velocity of the satellite.

When a satellite has the same periodic time as that of the earth, it will appear stationary when viewed from the earth's surface. Such satellites are said to be in a **parking orbit.** They are widely used in weather forecasting and in telecommunication.

27.7.3: Speed governor



As the shaft rotates, the masses also rotate with increasing angular velocity. Thus the angle θ enlarges. The collar is then pulled upwards by the arms which in turn pulls the lever up. The lever is connected to the fuel/steam valve which regulates the flow of fuel or steam which in turn controls the speed of the engine.

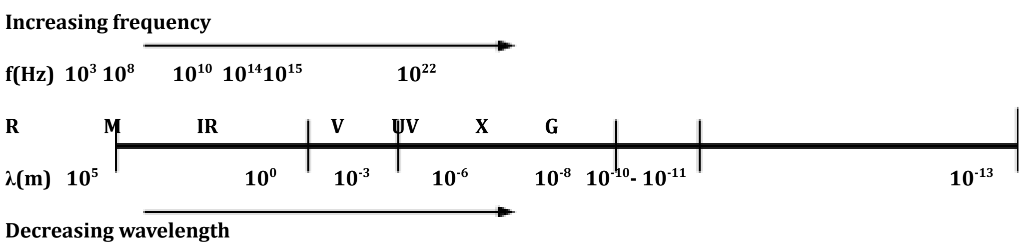
TOPIC 28: ELECTROMAGNETIC SPECTRUM

28.1: Introduction

We have seen that waves can be categorized either electromagnetic or mechanical in nature. Electromagnetic waves are waves resulting from the interaction of oscillating electric and magnetic fields. They include visible light, radio waves, x-rays, infra red, ultraviolet, microwaves and gamma radiations.

When these waves are arranged in a certain pattern e.g in the order of increasing frequency or wavelength then we get an **electromagnetic spectrum**.

28.2: The electromagnetic spectrum



Where **R-** Radio waves

M- Microwaves

IR- Infra red

V- Visible light

UV- Ultraviolet

X- X-rays

G- Gamma radiation

[**Hint:** Roast(R) maize (M) is (IR) a very (V) unusual (UV) x-mass(X) gift (G)**].**

28.3: Properties of electromagnetic waves

The following properties are common to all electromagnetic waves:

- ✔ Travel in a vacuum with a speed of $3.0x10^8$m/s.
- ✔ Do not require material medium for their propagation.
- ✔ Transverse in nature.
- ✔ Posses and transfer energy. The amount of energy possessed by an electromagnetic wave of frequency f is expressed as **E= hf**, where h is Plank's constant and is equal to $6.63x10^{-34}$Js. The wave equation **c=fλ**also apply for electromagnetic waves.
- ✔ Carry no charge (not charged) and are not deflected by a magnetic or electric field.
- ✔ Undergo reflection, refraction, diffraction, interference and polarization effects.
- ✔ Can be emitted, transmitted and absorbed by matter.

28.4: Production, detection and applications of electromagnetic radiations

The table below summarizes the production, detection and applications of the various electromagnetic radiations:

Radiation	**Production**	**Detection**	**Application**
Radio waves	From oscillating electrical circuits.	Antennae (aerials), diodes, earphones.	In telecommunication- radio broadcast, TV and satellite communication, cellular telephone, radar and navigation equipments etc.
Microwaves	From special vacuum tubes called magnetrons within microwave ovens.	Crystal detectors, solid state diodes, antennae.	Cooking in microwave cookers. In communication- mobile phones. In speed cameras.
Infra red	From thermal vibration of atoms in very hot bodiese.g the sun.	Thermopile, bolometer, thermometer, photographic film. Heating effect on the skin.	Burglar alarms, in military night vision missiles, cooking, heating and drying of grains, in green housing, in remote controls for TVs and VCD/DVDs, in photography.
Visible light	Sun, hot objects, lamps, laser beams.	The eye, photographic film, photocell.	Vision (sight), photography, photosynthesis and optical fibre. Laser beams used in laser printers, weapon aiming systems, CD players.
Ultraviolet	The sun, sparks, mercury vapour lamps.	Photographic film, photocells, fluorescent materials e.g quinine sulphate.	Detection of forgeries, skin treatment and killing of bacteria, spectroscopy and mineral analysis, making of clothes and a source of vitamin D.
X-rays	In X-ray tubes	Fluorescent screen and photographic film.	Radiography (identification of internal body structures e.g bones), cancer therapy, crystallography (study of crystal structure), pest and germ control and airport security ckecks.
Gamma radiation	Emitted by radioactive substances.	Radiation detectors e.g GM tube.	Sterilizing medical equipment and food. Killing of cancer cells and other malignant growths. Pest control. Detection of flaws in metals.

28.5: Dangers of electromagnetic waves

1. Some of the electromagnetic waves like X-rays, ultraviolet and gamma radiation posses a lot of energy. Therefore when exposed to the body in large doses, they can damage the body cells, cause skin burns or affect the eyes. Similarly, radio waves can cause cancer, leukemia among other disorders.
2. Nuclear reactor explosions may lead to losses of lives.

Example 28.1

1) A radio transmitter produces waves of frequency 1.0×10^8Hz. Calculate the wavelength of the signal.

$c=f\lambda$

$3.0 \times 10^8 m/s = 1.0 \times 10^8 \times \lambda$

$\lambda=3.0m$

2) An X-ray machine produces a radiation of wavelength 1.0×10^{-11}m. calculate:

a) The frequency of the radiation.

$c=f\lambda$

$3.0 \times 10^8 m/s = f \times 1.0 \times 10^{-11} m$

$f=3.0 \times 10^{19} Hz$

b) The energy content of the radiation. Take $h=6.63 \times 10^{-34}$Js.

$E=hf=6.63 \times 10^{-34} Js \times 3.0 \times 10^{19} Hz$

$=1.989 \times 10^{-14} J.$

3) Arrange the following radiations in order of increasing wavelength: infra-red, blue light, UV light, radio waves and X-rays.

X-rays, UV light, blue light, infra-red and radio waves.

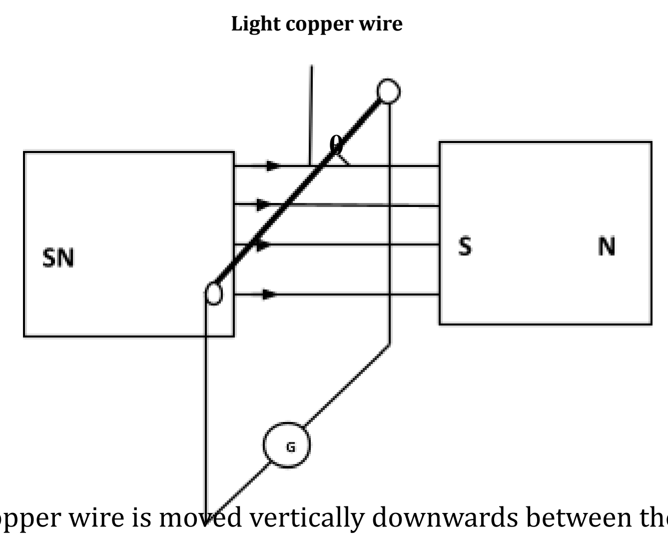
TOPIC 29: ELECTROMAGNETIC INDUCTION

29.1: Introduction

When a conductor moves within a magnetic field at an angle greater than zero, current is produced in the conductor which can be shown by connecting a galvanometer in series with the conductor. This method of generating electricity is called **electromagnetic induction**. It was first discovered by Michael Faraday about the year 1831.

Electromagnetic induction has been widely used to produce in large scale electrical energy in power stations.

29.2: Factors affecting the size of the induced electromotive force and Faraday's law



When the copper wire is moved vertically downwards between the poles of the magnet, the galvanometer is observed to deflect. However, the direction of deflection changes when the wire is now moved vertically upwards. When the conductor is kept stationary between the poles of the magnet, no deflection occurs. Similarly when the wire is placed parallel to the magnetic field, no deflection is observed.

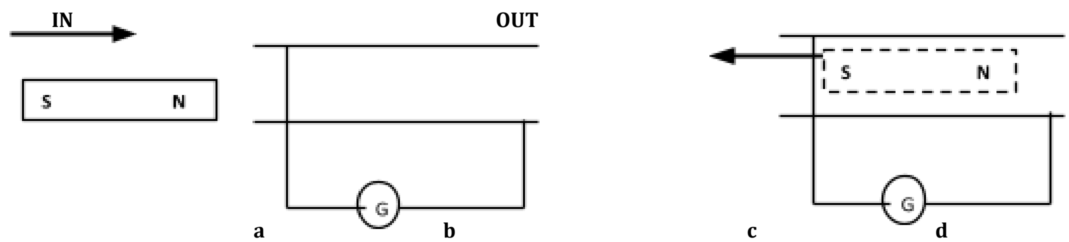
A deflection of the galvanometer indicates presence of induced electromotive force while absence of deflection indicates no induced electromotive force.

The deflection is maximum when the **angle between the wire and the field is 90^0, a stronger magnet is used and when the wire is moved very swiftly (at a high speed**).

These factors can be summed up in Faraday's law which state: **the magnitude of the induced emf is directly proportional to the rate of change of the magnetic flux linkage**. Magnetic flux linkage refers to the number of magnetic field lines cut by the conductor per unit area.

29.3: Lenz's law

Electromotive induction also occurs when a magnet is moved to and back within a solenoid as shown below:

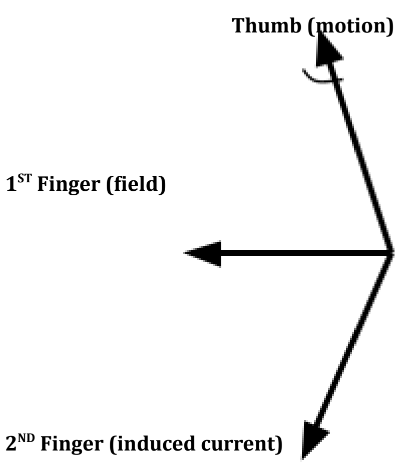


When the magnet is pushed into the solenoid, the galvanometer is observed to deflect same to when it is brought out but in opposite directions. However, when the magnet is kept stationary in the solenoid no deflection occurs.

Specifically, when the north pole of the magnet is brought into the solenoid the galvanometer deflects towards the left showing that current flows from **b to a** but deflects towards the right when the magnet is moved away from the solenoid showing that current flows from **c to d**.

These observations are summarized in Lenz's law which state: **the induced current flows in such away to oppose the change causing it**. It is based on the principle of conservation of energy i.e the mechanical energy of the moving magnet is converted to electrical energy in the form of the induced current.

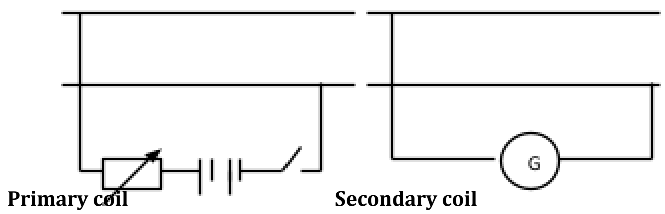
The direction of the induced current can be predicted by Fleming's right hand rule: if the thumb, first and second fingers of the right hand are held mutually at right angles to each other with the **F**irst finger pointing the direction of the **F**ield, thumb pointing the direction of the **M**otion then the se**C**ond finger points in the direction of the **C**urrent. It is also called the Dynamo rule.



29.4: Mutual induction

Mutual induction occurs when a varying current in one coil induces current in another close coil. The first coil in which current flows is called the **primary** coil while the second coil in which current is induced is called the **secondary** coil.

The varying current in the primary coil produces a magnetic field which links with the secondary coil inducing current in it.



When the switch is closed, current in the primary coil increases from zero to maximum. As a result, the magnetic flux linking up with turns of the secondary coil also increases from zero to maximum. This changing magnetic flux induces current in the secondary coil which makes the galvanometer to deflect. Once the current has reached maximum value, there will be no further increase in the magnetic flux and the pointer goes back to zero.

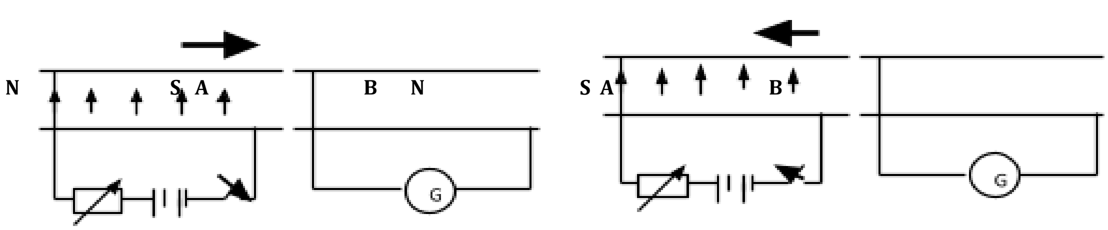
When the switch is open, the current falls from maximum to zero within a very short time. This implies that the magnetic flux of the primary coil takes a very short time to change. The shorter the time the higher the induced current and thus a larger deflection. Hence more current is induced during switching off than during switching on.

The magnetic flux of the primary coil linking up with the secondary coil can be varied by:

- Switching the current on and off.
- Varying the current in the primary coil using a rheostat.
- Applying an alternating current.

The direction of the induced current can be predicted applying the **Right-hand grip rule** and **Lenz's law** simultaneously. When doing so, the primary coil is treated as if it were a bar magnet moving into the secondary coil during switching on and as a bar magnet moving away from the secondary coil during switching off.

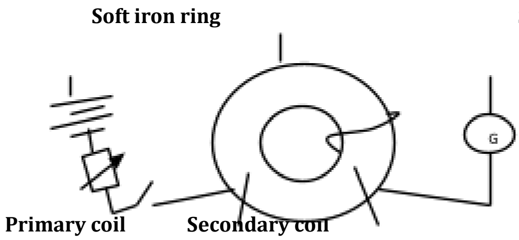
South pole approaching **South pole receding**

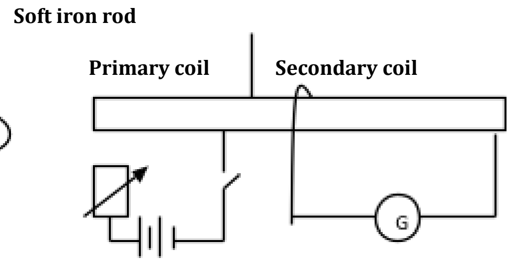


Thus by Lenz' law, during switching on the end A attains a South pole and B a North pole while during switching off the end A attains a North pole and B a South pole such that in each case the effect of the primary field is being opposed by that of the secondary coil.

The induced emf and current in the secondary coil can be increased by:

- Increasing the number of turns in the secondary coil so that many turns can link up with the magnetic field of the primary coil.
- Winding both the primary and secondary coils on a soft iron ring or soft iron rod. This will help to concentrate the magnetic field lines within the secondary coil. Soft iron is also easier to magnetize and demagnetize.

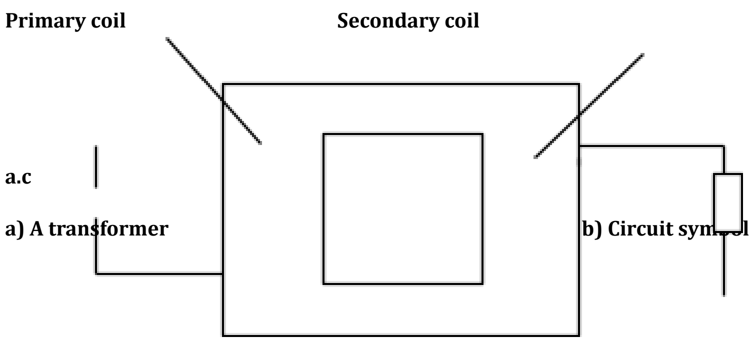




29.5: Applications of electromagnetic induction

29.5.1: Transformer

A transformer is a device that transfers electrical energy from one circuit to another by mutual induction. It comprise of two coils; the primary and secondary coils. An alternating current is fed into the primary coil whose magnetic flux links up with the secondary coil inducing current in it. Both the primary and the secondary coils are wound on a common soft iron core to enhance its effectiveness.



a) A transformer

b) Circuit symbol of a transformer

For any transformer, the ratio of the number of turns of the secondary coil $\mathbf{n_s}$to the number of turns of the primary coil $\mathbf{n_p}$ is called the turn's ratio;

i.e turn's ratio= n_s/n_p.

For an ideal transformer (100% efficient);
Power fed into the primary coil (power input)=power generated at the secondary coil (power output).
i.e$\mathbf{V_pI_p= V_SI_S}$.
Rearranging the above equation, we obtain:
$\mathbf{V_S/V_p= I_P/I_S}$.
Hence when the voltage is stepped up the current is stepped down.
Therefore, for an ideal transformer:
n_s/n_p $\mathbf{=V_s/V_p= I_P/I_S}$.

The efficiency of a transformer= {power output/power input}x100.
The equation $\mathbf{n_s/n_p = V_s/V_p}$ is referred to as the transformer equation.
There are two types of transformers namely step-down and step up transformer.

a) Step down transformer

In a step-down transformer, there are more turns on the primary coil than the secondary coil ($n_p > n_s$. thus the turn's ratio is less than one (1).

A step down transformer steps down voltage but steps up current. Hence the input voltage is less than the output voltage.

b) Step up transformer

A step up transformer steps up voltage but steps down current. In a step up transformer, there are more turns on the secondary coil than the primary coil. Thus the turn's ratio is more than one. The output voltage is greater than the input voltage.

Example 29.1

1. Calculate the number of turns of the secondary coil of a step down transformer which would enable a 12V bulb to be used with a 240V ac supply if there are 480 turns in the primary coil.

$n_s/n_p = V_s/V_p$

$n_s/480 = 12V/240V$

$n_s = (480x12)/240 = 24$turns

2. What current will flow in the secondary coil when the primary current is 0.5A if the voltage in the primary coil is 240V and that in the secondary coil is 48V? Assume the energy loss is negligible.

$\mathbf{V_S/V_p = I_P/I_S}$

$48V/240V = 0.5A/I_S$

$I_S = (0.5x240)/48 = 2.5A$

3. A student designed a transformer to supply a current of 10A at a potential difference of 60V to an electric motor from an a.c supply of 240V. If the efficiency of the transformer is 80%, calculate:

a) The power supplied to the transformer.

Efficiency=(power output/ power input)x100

80=[(60x10)/power input]x100

Power input= (60x10x100)/80

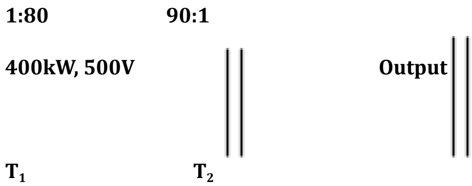
=750W

b) The current in the primary coil

Power input=$V_P x I_P$

I_P=750W/240V =3.125A

4. A model of a transformer system consists of a power generator, a step up transformer, transmission cables and a step down transformer.



Given that the total resistance of the cables is 200Ω, transformer T_1 is 100% efficient and T_2 is 95% efficient, calculate:

a) The current through the primary coil of T_1.

Power input=V_P I_P

100,000W=500VxI_P

I_P=100,000/500 =800A

b) The voltage across the secondary coil of T_1.

$n_s/n_p=V_S /V_P$

80/1= V_S/500V

V_S= 500x80=40000V

c) The voltage across the primary coil of T_2.

V_P I_P=V_S I_S

400,000=40000VxI_S

I_S=400,000/40,000 =10A

This is the current flowing through the secondary coil of T_1.

Hence the voltage dropped across the cables (lost through resistance)= IR

V= 10x200= 2000V

Therefore the voltage across the primary coil of T_2=40,000-2000

=38,000V.

d) The maximum power output of the transformer T_2.

Efficiency=(power output/power input)x100

95= (power output/38000x10) x100

Power output= [95x38000x10]/100

=361,000V

- **Energy losses in transformers**

There are four main causes of energy losses in transformers:

i. **Flux leakage**

When part of the magnetic flux of the primary coil fails to reach the secondary coil, it is referred to as magnetic flux leakage.

In order to minimize energy loss through flux leakage can be reduced by winding the primary and secondary coils next to each on a common core. Alternatively, the secondary coil can be wound on top of the primary coil.

ii. **Resistance**

As current flows through the coils, heat is generated due to the resistance of the coils. The electrical power loss as a result of resistance is given by= $\mathbf{I^2R}$.

Energy loss through resistance can be minimized by using thicker cooper wires.

iii. **Eddy currents**

As the current alternates the magnetic flux also keeps alternating in the soft iron core producing eddy currents. These currents are sufficient enough to generate heat within the core.

The energy loss can be minimized by laminating the core i.e. using thin sheets of soft iron plates insulated from each other.

iv. **Hysteresis loss**

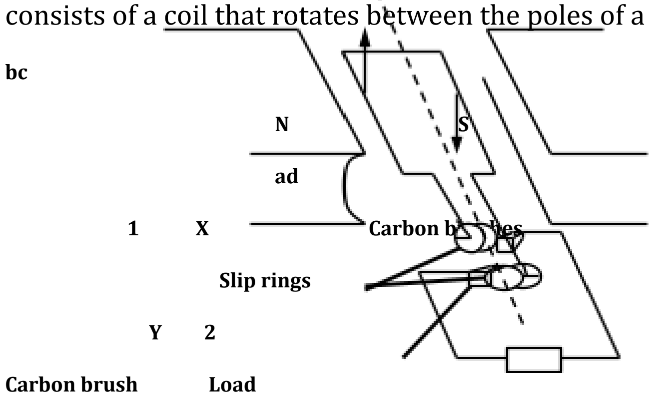
The process of magnetization and demagnetization any time current reverses do generate heat within the core. Energy loss in this way is referred to as hysteresis loss.

Hysteresis loss can be minimized by using soft iron core which is easier to magnetize and demagnetize.

Note that despite the above measures, some of significant heat is still generated within the transformers. This is further cooled using oil.

29.5.2: Alternating current (ac) generator

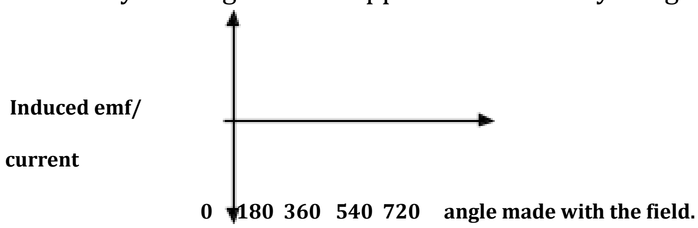
It is also called the alternator. A generator is a device that converts mechanical energy into electrical energy. It consists of a coil that rotates between the poles of a strong magnet.



Suppose the coil is rotated in the clockwise direction as shown above, then the sections **ab**and **cd** cut through the magnetic field between the poles of the magnet inducing current in the coil. Applying Fleming's right hand rule on the section **ab,** the induced current flows in the direction **abcd**and through the load via the slip ring 2 and carbon brush Y.

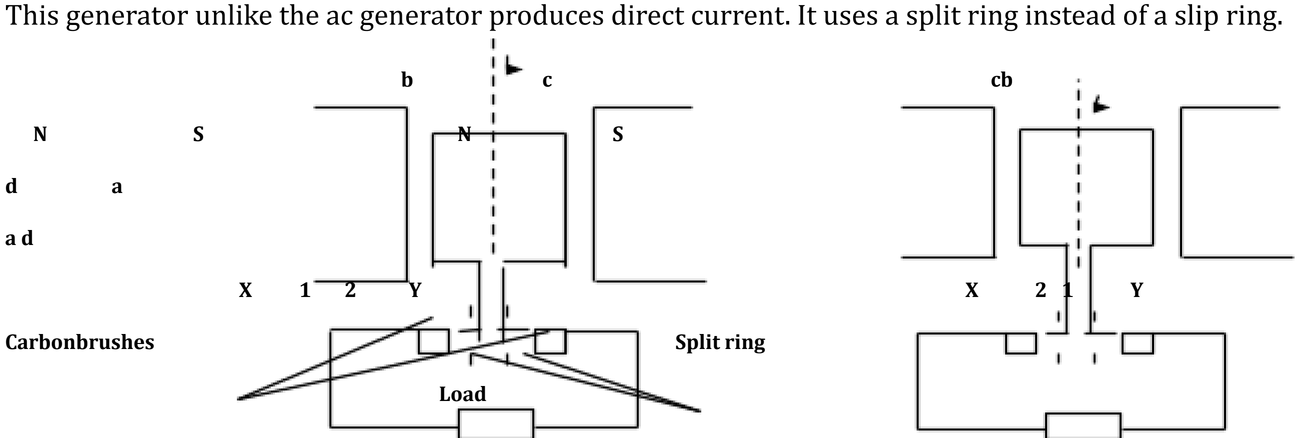
The induced current is maximum when the coil is horizontal and reduces to zero when the coil is at the vertical position. At the vertical position, the coil is parallel to the magnetic field and thus no current is induced in the coil.Beyond the vertical position, the induced current increases from zero to maximum when the coil is once again at the horizontal position.

Applying Fleming's right hand rule again, the induced current flows in the direction **dcba**and through the load via the slip ring 1 and carbon brush X. Hence in second half cycle, current direction is reversed. The induced emf or current by the ac generator appears as shown by the graph below:



29.5.3: Direct current (dc) generator

This generator unlike the ac generator produces direct current. It uses a split ring instead of a slip ring.



(a) (b)

As the coil rotates from the horizontal position to the vertical position, the induced current reduces from maximum to zero. When the coil is rotated in the clockwise direction as shown in figure (a), the induced current flows in the direction **abcd**and through the load via the half split ring 2 and carbon brush Y.

At the vertical position the carbon brushes touch the gaps between the commutators (half split rings). When the coil goes past the vertical position, the half split rings automatically exchange the carbon brushes as shown in figure (b) above. Applying Fleming's right hand rule again, the induced current flows in the direction **dcba**and through the load via the half split ring 1 and carbon brush Y. thus the direction of current through the load is in one direction only hence direct current generator.

The graph of induced current/emf by a dc generator appears as shown below:

0 180 360 540 720 angle made with the field

Note that the magnitude of the induced current and electromotive force produced by both the ac and dc generators can be increased in the following ways:

- ✔ Using a stronger magnet.
- ✔ Increasing the number of turns of the secondary coil.
- ✔ Increasing the speed of rotation of the coil.
- ✔ Winding the coil on a soft iron core.

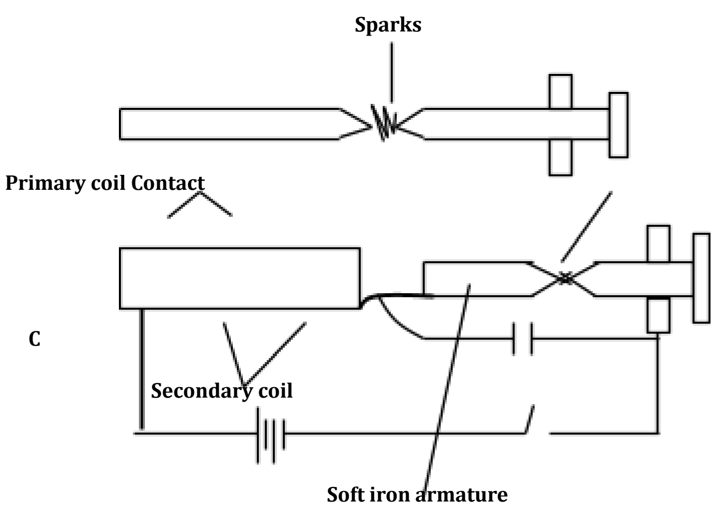
In some generators like the bicycle dynamo, the coil is kept stationary while the magnet rotates.

29.5.4: Induction coil

An induction coil consists of a primary coil of fewer turns and a secondary coil of many turns, both wound on a soft iron core. When the switch is closed, the soft iron core becomes magnetized and attracts the soft iron armature. This breaks the contact and current stops flowing. The core is thus demagnetized. The armature is released and the contact is remade. The process is repeated as long as the switch is closed.

The changing magnetic flux during magnetization and demagnetization produces an induced emf and current in the secondary coil. The induced current in the secondary coil produces sparks as current flows through the air across the gaps at the ends of the secondary coil as the shown in the figure below:

..

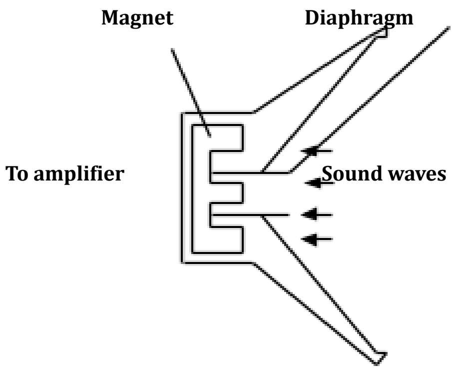


The capacitor is used to store charge and thus reduce sparking.

The sparks produced can be used to ignite an air-petrol mixture such as in the car ignition system.

29.5.5: The moving coil microphone

A moving coil loudspeaker converts sound energy into electrical energy.



When a person speaks into the microphone, the sound waves set the diaphragm into vibration. This makes the coil to move forward and back cutting through the magnetic field inducing current in the coil. The induced current flows to the amplifier for amplification before it is relayed to the loudspeaker where it is converted back to sound energy.

TOPIC 30: MAINS ELECTRICITY

30.1: Introduction

Mains electricity refers to the electrical power supplied to households. The power is generated at the power station and then transmitted to the consumers either through overhead transmission lines or underground cables.

Some of the sources of mains electricity include water in high dams, geothermal, wind, solar energy, coal and diesel engine generators, nuclear energy and tidal waves.

Note that the choice of the source of mains electricity to use is dependent on its availability and abundance as well as the implications it has on the environment.

In Kenya, the most utilized source of mains electricity is hydroelectricity. Today, there is also increased usage of solar energy, geothermal energy, wind energy and coal.

30.2: Electrical power transmission

This is process by which electrical power is relayed from the generation plant to the consumers at their homes, institutions, schools, industries, factories etc. In Kenya, electrical power is distributed using the national grid system. The national grid system is a network of cables connecting at a common point all the power generation plants and then distributed to the consumers. This way it ensures availability of power even when one of the stations is shut down.

Before power is fed into the national grid system, it is stepped up i.e. voltage is stepped up but current is stepped down. Most power stations generate between 11kV and 25kV which is stepped up to between 132kV and 400kV. At the consumer end, a step down transformer is used to step down the voltage to about 11kV at a substation. However, this value is still large to be used the way it is. The power is then distributed by cables to another step down transformer which is situated just near the consumer. This transformer further steps down the voltage to a consumable value of 240V for domestic use or 415V for factories.

High voltage transmission has the following three benefits:

- ✔ Power loss due to resistance of the cables is minimal.
- ✔ Thin cables could be used.
- ✔ Reduced cost in acquiring cables and poles.

However, high voltage transmission has some limitations. These include:

- High risks of electric shock in case the poles collapse or the cables hang very low.
- High risks of fire outbreak on nearby buildings and vegetation when the cables hang too low or when the cables touch each other.
- Harmful effects of strong magnetic fields due to the current flowing in the cables.

Electrical power is generated and distributed as alternating current. This is because of the following reasons:

i. Transformers work on alternating current/voltage only.

ii. It is easier to step up or down.

iii. Alternating current is easier to rectify.

Example 30.1

1. Some length of a power line has a resistance of 10Ω and is transmitting 11kV and a current of 1A. If the voltage is stepped up to 160kV by a transformer, calculate:

    a) The initial power loss

Power loss= $I^2R = 1^2 x10$

$= 10W$

    b) The power loss after step up

$V_P I_P = V_S I_S$

$10kV \times 1A = 160kV \times I_S$

$I_S = (10/160) = 0.06875A$

Hence, power loss= $0.06875^2 \times 10$

= 0.048W

2. A power station produces 50kW at 240V. The power is transmitted through cables with a resistance of 0.4Ω. Calculate the percentage power loss during transmission.

Power=VI

50,000W= 240VxI

I=50,000/240 =208.33A

Hence, power loss= $208.33^2 \times 0.4$

=17,360.556W

%power loss= (17360.556/50000)x100 =34.72%

30.3: Domestic wiring

From the step down transformer near the consumer, power is transmitted by two cables; the live and neutral wire to the consumer's meter which measures and registers the amount of power consumed. The live wire is at full potential of 240V while the neutral wire is at zero potential since it has been earthed at the sub-station. From the meter, electricity enters the fuse box which comprises of the following:

- ✔ Main switch- controls all the live and neutral wires simultaneously. It is normally useful during repairs.
- ✔ Live busbar- it is a brass bar on which all the live wires of all the circuits have been connected. Each live wire is connected to the live busbar through a fuse.
- ✔ Neutral busbar- it is a brass bar to which all neutral wires of all the circuits have been connected.
- ✔ Earth terminal- it is used to earth the circuit. This can be done by burying a thick copper wire deep underground or through a metallic water piping.

A fuse is a thin wire with very low melting point such that if it is overheated it melts and the circuit gets disconnected. This way it is used to safeguard electrical appliances. Fuses are rated in amperes. Normally, the fuse rating is slightly above the maximum current requirement of the appliance. All fuses must be connected along the live wire.

In domestic wiring, there are three important types of wires commonly used namely live wire, neutral wire and earth wire.

- ✔ Live wire- it transmits alternating current from the source to the appliance or plug. It is normally red or brown in colour.

- ✔ Neutral wire- it is the return wire i.e. it returns the current back to the source completing the circuit. It is usually at zero potential. It is normally blue or black in colour.
- ✔ Earth wire- it earths the circuit. It is normally green or yellow in colour

However, at times the fuse may melt off and thus fail to serve its rightful purpose. Some causes that can lead to melting off of the fuse include:

i. Short circuiting when bare conductors touch each other.

ii. Overloading the circuit with more appliances than the fuse can accommodate.

iii. Using a fuse of lower rating than the current requirement of the appliance.

iv. Using a faulty fuse whose wire could have been oxidized.

Example 30.2

1. An electric cooker has an oven rated 3kW, a grill rated 2kW and two rings each rated 500W. The cooker operates from 240V mains. Would a 30A fuse be suitable for the cooker assuming that all the parts are switched on?

I_{oven}=3000/240 =12.5A

I_{grill}=2000/240 =8.33A

$2I_{rings}$=(2x500)/240 =4.17A

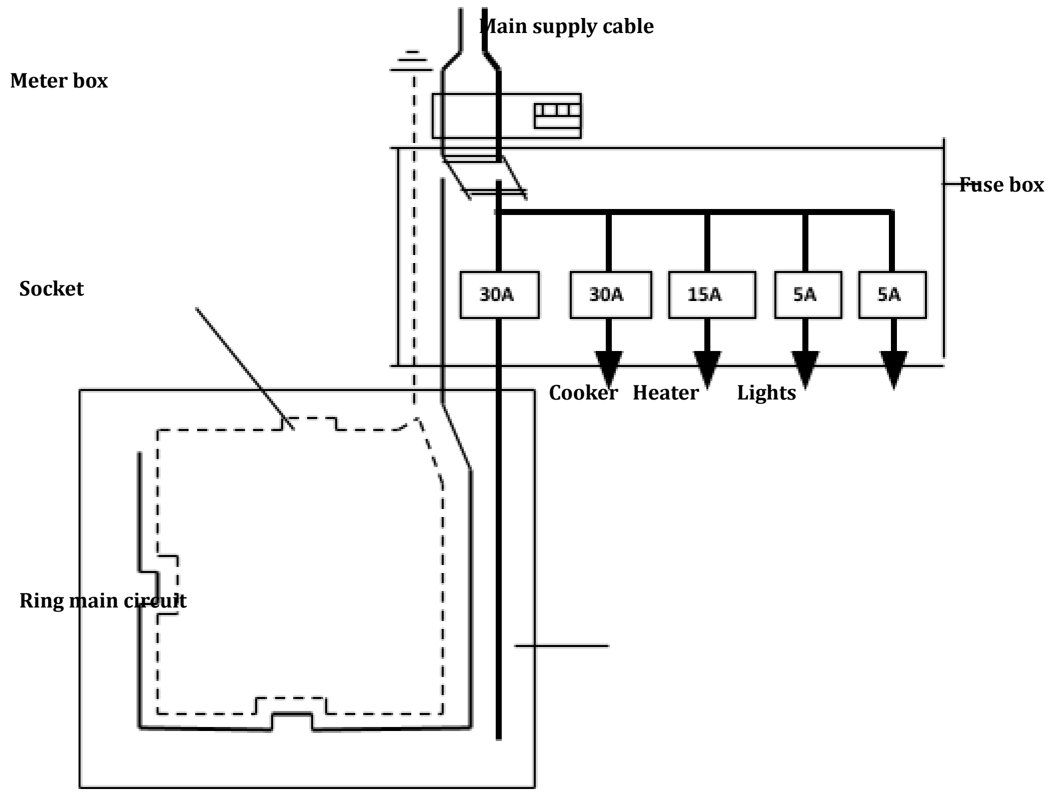
Total cooker current =12.5+8.33+4.17

=25.0A

With a fuse rated 30A, it is suitable.

Note that a circuit breaker can also be used to serve the same purpose as the fuse. It should be noted that a circuit breaker is better than a fuse since for a fuse once it has blown off it must be replaced while for a circuit breaker, it does not need to be replaced. Instead, the strip just needs time to cool off and then the circuit will be complete once again.

The diagram below shows a typical house wiring system:



30.3.1: The lighting circuit

This is the circuit that controls all the lamps within the house. Lamps are always connected in parallel so that they are operated from the same mains voltage. This also ensures that the other lamps continue to work when one is faulty. The lighting circuit uses very low current and therefore thin wires are usable. This is also the reason why the lighting circuit needs not to have the earth wire. It also has a low rated fuse, mostly 5A fuse.

A two-way switch circuit

In this circuit, a lamp is operated by two switches i.e. one can put the lamp on and the off and vice versa. Such a circuit is very convenient for lighting a staircase or corridors.

Note that all switches are connected along the live wire.

30.3.2: The cooker circuit

The electric cooker and water heater both use a lot more current and therefore require a high rated fuse. In most cases, they are normally supplied with their own circuits. Such circuits are usually earthed and use thicker wires.

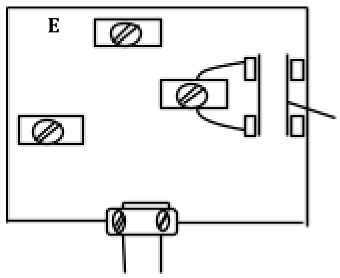
30.3.3: The ring main circuit

This circuit supplies power to the sockets around the house. It also uses more current compared to the lighting circuit.

Sockets are used with plugs. There are two types of plugs namely a two-pin plug and a three-pin plug. A two-pin plug has only the live pin and the neutral pin while a three-pin plug has in addition to the above two an earth pin which is normally longer than the other two pins. This pin has two roles:

i. It is used to open the blinds (socket) for the other two pins.

ii. It is used to leak out the charges which could have developed in the socket.



Some plugs have a fuse inserted in them along the live wire to safeguard appliances in case of excessive current.

30.4: The cost of electrical energy

The cost of electrical energy consumed depends on the power rating of the appliances used and the duration of use. The electrical energy used is measured by the meter in kilowatt-hours (kWh). It is given by the product of the power in kilowatts and the time in hours;

Electrical energy consumed= power (kW) x time (h).

One kilowatt-hour is also referred to as one unit;

1kWh= 1unit

The total cost of electrical energy consumed by a household is given by the product of the number of units consumed and the charges per unit;

Total cost= number of units used x cost per unit.

Example 30.3

1. A six bulb arrangement in a house runs for 8hours every night for 5days. If each bulb is rated 100W and the cost of electrical energy is sh. 2.60 per unit, how much will the owner of the house pay at the end of the five days?

Electrical energy consumed= (6x0.1x8) x5= 24kWh

Total cost= 24kWh x sh 2.60= ksh. 62.40

2. An electric cooker has an oven rated 3kW, a grill rated 2kW and two rings each rated 500W. The cooker operates from a 240V mains. What is the cost of operating all the parts for 30minutes if electricity costs Sh 1.50 per unit?

Total units consumed= [3+2+(2x0.5)]x30/60

=3kWh

Total cost= 3kWh x sh 1.50

= Ksh 4.50

TOPIC 31: CATHODE RAYS AND CATHODE RAY TUBE

31.1: Introduction

When a metal surface is heated, the electrons gain energy and become excited. At very high temperatures, the electrons may break off from the force of attraction of the nuclei. When heat is used to extract electrons from the surface of a metal, it is referred to as **thermionic emission**.

Cathode rays are streams of fast moving electrons emitted from the surface of a heated cathode inside a vacuum.

31.2: Production of Cathode Rays.

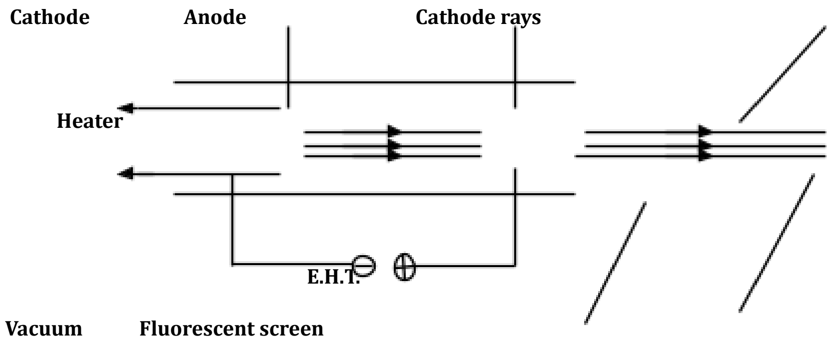


Fig.1 Cathode Ray Tube

Cathode rays are produced in a cathode ray tube. The cathode is heated by the heater emitting electrons through thermionic emission. Note that the cathode rays are streams of negatively charged particles (electrons). Thus once emitted at the cathode, the electrons will be attracted by the anode which is at a positive potential. Hence the role of the anode is to accelerate the electrons towards the screen. The anode is connected to an extra high tension (EHT) source.

The tube is evacuated. This is to prevent the electrons from interacting with any particles before reaching the screen. The screen is coated using a fluorescent material that glows when struck by the electrons.

31.3: Properties of Cathode Rays

1. They travel in straight lines. When an opaque object is placed along the path of the rays, a sharp shadow of the object is formed on the screen.

2. They are charged. Hence they are deflected by both magnetic and electric fields.

3. They posses kinetic energy.

4. They can cause certain substances e.g zinc sulphide screen to glow or fluoresce.

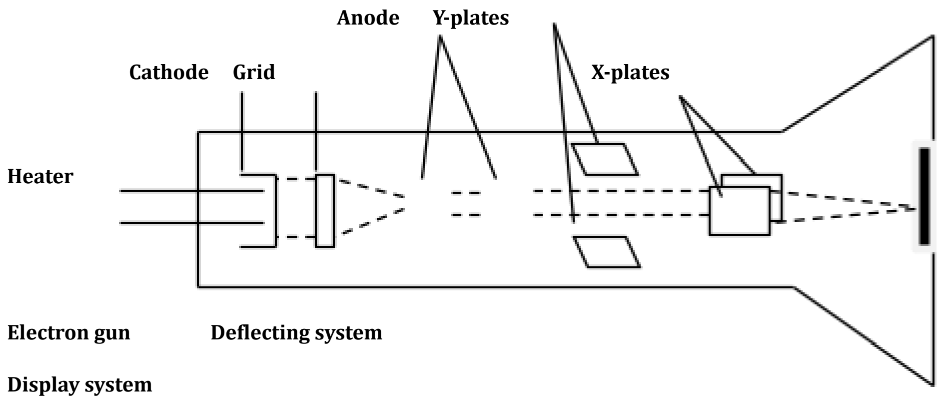
5. They can produce X-rays when they are suddenly stopped by a metal target.

31.4: The Cathode Ray Oscilloscope (CRO)

This is an electrical instrument developed from the cathode ray tube and which can be used to display and analyze waveforms. It can display both alternating current and direct current waveforms. Furthermore, it can be used to measure voltages that vary over time.

A cathode ray oscilloscope has three main components:

- ✔ The electron gun.
- ✔ The deflecting system.
- ✔ The display system.



a) The electron gun

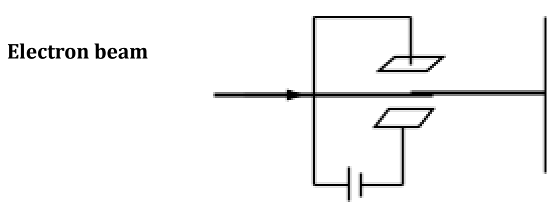
It consists of three parts namely the cathode, grid and anode. The cathode emits electrons through thermionic emission. The grid concentrates the electrons into a tight beam. It is connected to the negative terminal of the EHT and thus it is at a negative potential. When the negative voltage of the grid is raised, fewer electrons will move towards the screen and thus the spot will be less bright. However, when the grid voltage is lowered, more electrons will move towards the screen and thus the spot will be brighter. In general, the grid controls the rate of flow of electrons to the screen i.e. intensity.

The anode on the other side is at a positive potential and is used to accelerate the emitted electrons towards the screen. It also focuses the electrons to a point on the screen.

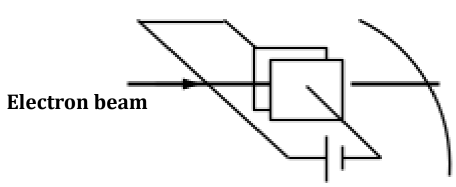
b) The deflecting system

This system places the electron beam on the screen.it comprise of two pairs of parallel plates namely the Y-plates and X-plates.

The Y-plates are responsible for the vertical deflection. When the upper plate is at a positive potential for instance, the beam is deflected upwards while if the lower plate is now at a positive potential, the beam is deflected downwards. However, when both plates are at a zero potential the beam will pass undeflected.



The X-plates are responsible for the horizontal deflection of the electron beam.



There is no deflection when the potential difference across the plates is zero but deflects towards the plate at a positive potential when connected to a source of voltage as shown on the figure above.

If an alternating voltage is applied simultaneously to both the Y and X-plates, then the spot on the screen would oscillate up and down and at the same time move across the screen from left to right tracing a wave on the screen. When the spot reaches the extreme end it flies back to the starting point and process is repeated.

The speed with which the spot moves on the screen can be adjusted by the time base knob.

31.5: Uses of the CRO

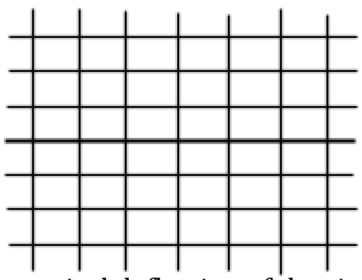
i. **Used as a voltmeter.**

The time base is switched off while the voltage to be measured is fed through the Y-gain of the CRO. The applied voltage displaces the spot vertically on the screen. The Y-gain control knob can be used to amplify the display on the screen by setting it to a certain value. This is referred to as the sensitivity of the CRO.

Hence the corresponding voltage to the signal on the screen is expressed as;

Voltage (V)= vertical deflection (cm) x Y-gain setting or sensitivity (volts/cm)

Consider the waveform shown in the figure below:



The maximum vertical deflection of the signal is 3cm or 3divisions on either side. Suppose the Y-gain setting is 50V/cm;

Then the peak voltage represented by the signal = vertical displacement (cm) **x** Y-gain setting (V/cm).

=3cm x50V/cm =150V

When used as a voltmeter, a CRO has the following advantages over the ordinary voltmeters:

- ✔ Can measure both direct and alternating voltages.
- ✔ Can measure very large voltages without being damaged.
- ✔ Does not take any current in the circuit since it has infinite resistance.
- ✔ It responds instantly. The pointer of ordinary voltmeters always swings about the correct value.

ii. **Used to measure frequency**

The signal whose frequency is to be measured is fed into the Y-gain and the time base is switched on and adjusted so that the waveform appears stationary on the screen. Suppose the time base setting is 10ms/cm, it implies that the wave takes 10ms to cover 1cm horizontally. This can be used to determine the time of one wave i.e period T.

Recall: frequency=1/ period T.

In the example above, suppose the time base is set at 40ms/cm;

Then, period T= number of divisions fitting one wave x time base setting

=4cm x (40/1000) s/cm

=0.16s

Hence frequency of the signal= 1/0.16 =6.25Hz

Other uses of the CRO include:

- ✔ Measurement of small time intervals.
- ✔ Measurement of amplitudes of direct and alternating voltages.
- ✔ Display of electrical signals whose variations can be put in the form of voltage.

31.6: The television tube

A television tube comprises of three electron guns, two sets of coils for deflection and a fluorescent screen. There are two types TV tubes namely the black and white tube and a coloured tube. The signal is fed into the television through the control grid. This varying incoming signal regulates the number of electrons being emitted by the electron gun at any instant. This in turn regulates the brightness of the spot on the screen.

The screen is coated using green, blue and red phosphors. When a red colour is required for instance, the red electron gun emits a red electron beam which strikes the red phosphor on the screen. The same happens for green and blue colours.

For white colour, all the three electron guns simultaneously fire electrons to the screen. However, for black colour, no beam is fired to the screen. The rest of the colours are just obtained from a combination of any two of the three colours.

The deflection of the beams is done by the two coils; one responsible for the vertical deflection and the other for the horizontal deflection. A current is fed into these coils producing a magnetic field. As the electrons pass between the coils, the resultant force on them causes them to deviate.

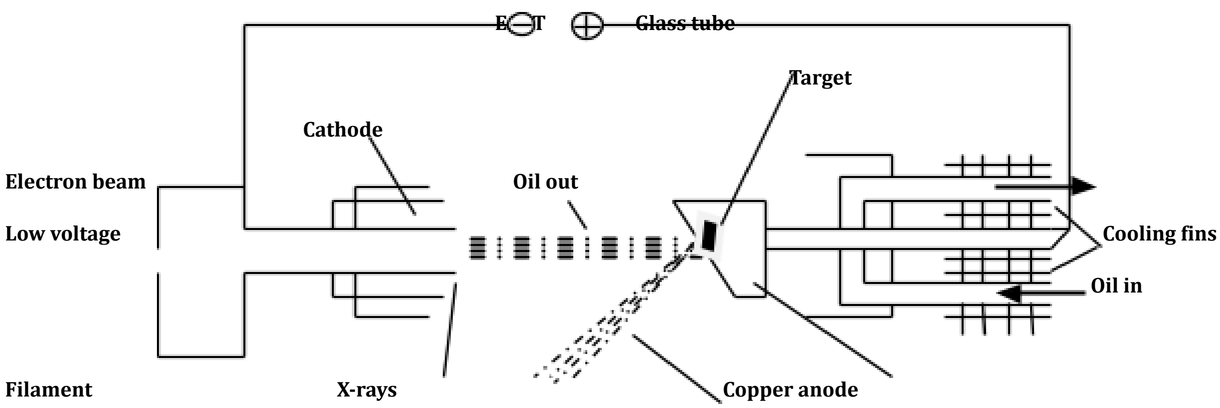
A magnetic field is preferred to an electric field in a TV tube because it gives a wider deflection on the screen and thus a shorter tube can be used.

TOPIC 32: X-RAYS

32.1: Production of x-rays

X-rays are produced when fast moving electrons are suddenly stopped by a metal target. At the time of their discovery by a German Physicist Ron6tgen, their nature was unknown and hence their name x-rays. Generally, x-rays are uncharged electromagnetic radiations of short wavelength and high penetrating ability (power).

X-rays are produced in an x-ray tube:



Current in the filament emits electrons at the cathode by thermionic emission. These electrons are then attracted towards the anode by the high potential difference that exists between the cathode and anode. On striking the target, the electrons transfer their kinetic energy to the metal target. About 99.5% of this energy is converted to heat at the target and only 0.5% of the energy is responsible for the production of x-ray radiations.

As such, the material of the target must be one that has a high melting point like molybdenum or tungsten. The anode should also be a good thermal conductor like copper so as to ensure efficient dissipation of heat.

Further cooling at the anode is enhanced by a circulation of oil around the anode and the presence of cooling fins. In some tubes, the target is made in such away to rotate so as to change the point of impact and thus reduce wear and tear.

The target is inclined at an angle to direct the x-rays out of the tube. The glass tube is also evacuated to **prevent interference with the electron beam before reaching the target.** The cathode is concave in shape **to focus the emitted electrons to the target**. The high potential difference is used to accelerate the emitted electrons towards the anode.

The x-ray tube is well shielded using lead which absorbs any stray rays thereby protecting the user.

32.2: Properties of X-rays

- Travel in straight lines with the speed of light in air; $3.0x10^8$m/s. When an opaque object like a bone is placed on the path of x-rays a sharp shadow of the object is formed on the screen.

- They carry no charge. Hence x-rays are not deflected by either magnetic fields or electric fields.

- Ionize air molecules on their paths by knocking off electrons in them.

- They cause certain substances and salts to fluoresce.

- They cause photographing emulsion, a property used in x-ray photography.

- They cause photoelectric effect when incident on the surface of some metals.

- They can readily penetrate matter. The degree of penetration depends on the density of the material and the quality of the x-rays.

- They obey the wave equation $v=f\lambda$.

- They undergo interference, reflection, refraction and diffraction effects.

32.3: Energy and Quality of x-rays

When an electron of charge **e** is accelerated by a voltage V applied across the tube, the electron gains an amount of energy equivalent to **eV** (electron volts). This energy is converted into kinetic energy of the electron;

i.e**eV=K.E**

$eV= ½mv^2$, where m- mass of the electron ($m=9.11x10^{-31}$kg) and v- the velocity of the electron.

Also, according to Plank's theory, the energy of any electromagnetic radiation x-rays included is given by;

Energy, E= hf, where h- is Plank's constant and f- is the frequency of the radiation.

Hence for x-rays; **$eV=½mv^2=hf=hv/\lambda$**.

Generally, most energetic x-rays are those with higher frequency or shorter wavelength while the least energetic x-rays are those with lower frequency or longer wavelength.

The energy of x-rays depends on the accelerating potential between the cathode and the anode. The higher the accelerating potential, the higher the energy of the electrons.Since it is the energy of the electrons that is converted into x-rays, the higher the energy of the electrons the higher the energy of the x-rays.

X-rays produced by high energetic electrons or high accelerating voltage are referred to as **hard x-rays**. They are high quality x-rays, have very high frequency and high penetrating power.

X-rays produced from low energy electrons or low accelerating voltage are called **soft x-rays**. They are low quality x-rays, have low energy content, low frequency and low penetrating power.

32.4: Intensity of X-rays

Intensity of x-rays refers to the number of x-rays produced per second. It depends on the number of electrons striking the target per second. This is controlled by the filament current. The higher the filament current the higher number of electrons emitted and hence the greater the intensity of the x-rays.

32.5: Detection of X-rays

X-rays can be detected by:

- Using a fluorescent screen. The screen glows when struck by the x-rays.

- Using a photographic plate. The plate is blackened when exposed to x-rays.

32.6: Uses of X-rays

In medicine

- Detection of fractures, displaced bones or other strange objects within the body.

- Destruction of cancerous growths and other malignant growths.

- Testing densities of bones.

- Detection of lungs with tuberculosis.

In industries

- Detection of flaws in metals and welded joints.

- Checking percentages of certain elements in an ore.

- For security checks in airports.

- To check the purity or genuineness of certain precious stones like gold, silver etc.

- To sterilize surgical equipment before packaging.

- Detection of leakages in water pipes.

In crystallography

- To study the crystal structure of substances.

32.7: Dangers of X-rays

Excessive exposure of living body tissues to x-rays may lead to damage or killing of the cells. X-rays can cause deep rooted burns, mutation and serious diseases.

These can be minimized by:

1. Limiting the exposure time of living tissues to x-rays.

2. X-ray sources should be well screened or shielded.

Example 32.1 (take $h=6.63\times10^{-34}$Js, $e=1.6\times10^{-19}$C, $m_e=9.11\times10^{-31}$kg and $v=3.0\times10^{8}$m/s)

1. Calculate the energy of x-rays whose frequency is $3x10^{16}$Hz.

$E=hf=6.63x10^{-34}Jsx3x10^{16}Hz$

$=1.989x10^{-17}J$

2. In an x-ray tube, an electron is accelerated by a potential difference of 1kV.

a) Determine the velocity of the electron as it is reaching the target.

$eV=\frac{1}{2}mv^2$

$1.6x10^{-19}Cx1000V=\frac{1}{2}\ x9.11x10^{-31}kg\ xv^2$

$v^2=3.5126x10^{14}$

v=

b) How much kinetic energy will the electron have acquired when it hits the target?

$eV=K.E=1.6x10^{-19}Cx1000V$

$=1.6x10^{-16}J$

3. Explain how you can increase:

a) Quality of x-rays.

By increasing the accelerating potential between the cathode and the anode.

b) Intensity of the x-rays.

By increasing the filament current so that more electrons are emitted per unit time.

4. An x-ray tube operates at 10kV and a current of 15mA. Calculate the number of electrons hitting the target per second.

I=ne

$15x10^{-3}A=nx1.6x10^{-19}C$

$n=9.375x10^{16}$electrons.

5. An x-ray tube operates at 20kV. What is the shortest wavelength in its x-ray beam?

$eV=hv/\lambda$

$1.6x10^{-19}x20000V=(6.63x10^{-34}Jsx3.0x10^{8})/\lambda$

$\lambda=6.2156x10^{-11}m$

6. State any differences between x-rays and cathode rays.

- X-rays are uncharged while cathode rays are charged.

- X-rays are produced in an x-ray tube while cathode rays are produced in a cathode ray tube.

TOPIC 33: PHOTOELECTRIC EFFECT

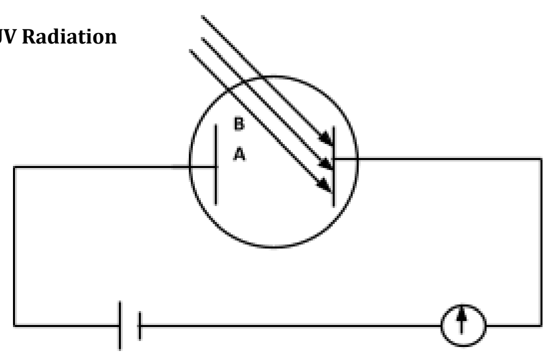
33.1: Photoelectric emission

We have already seen that when a metal surface is heated to a certain extent, electrons are dislodged. This is called thermionic emission. Similarly when a metal surface is irradiated using an electromagnetic radiation of a certain amount of energy, electrons are emitted. This process is called photoelectric emission or effect. The energy of the radiation is transferred to the electrons in the atoms of the metal. The electrons gain enough energy and get dislodged from the metal surface. These electrons are called photoelectrons.

Photoelectric emission can be shown by the following set-ups:

33.1.1: Using a galvanometer

UV Radiation

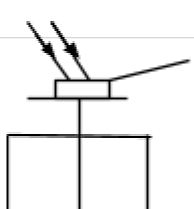


When the UV radiation is incident on the metal plate A, electrons are emitted which are then attracted towards plate B due to its positive potential. This completes the circuit and the galvanometer deflects.

However, when a glass barrier is placed along the path of the UV radiation no deflection will be observed as the glass cuts off the radiation from reaching the metal plate hence no photoelectrons are emitted.

33.1.2: Using a clean zinc plate and uncharged electroscope.

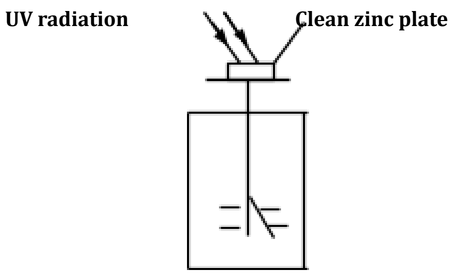
UV radiation



Clean zinc plate

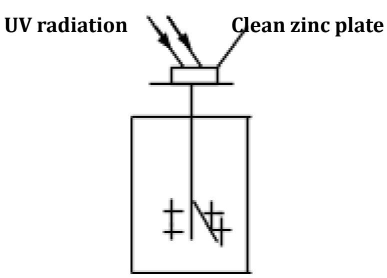
When UV radiation is incident on the clean zinc plate, electrons are dislodged from the zinc plate. The zinc plate thus loses electrons. Some electrons are then attracted from the plate and leaf of the electroscope towards the zinc plate leaving the electroscope positively charged. Hence the leaf of the electroscope diverges.

33.1.3: Using a clean zinc plate and a charged electroscope



The UV radiation dislodges electrons from the surface of the zinc plate leaving it with a deficit of electrons. It then attracts some electrons from the leaf of the electroscope. This in effect discharges the electroscope and the leaf divergence reduces with time.

However, when a positively charged electroscope is used, the UV radiation dislodges electrons which are immediately attracted back by the positive charges on the electroscope. Thus the leaf divergence remains unchanged.



33.2: The quantum theory and Einstein's equation

This theory was advanced by Max Plank. He says that electromagnetic radiations like light are propagated in small packets of energy called **quanta [singular- quantum**]. The amount of energy of a quantum is referred to as a **photon**.

According to Plank, the energy of a photon is directly proportional to the frequency of the radiation;

$E \alpha f$

Thus, $E=hf$: where h is Plank's constant [i.e $h=6.63\times10^{-34}$Js].

Since all electromagnetic radiations obey the equation $c=f\lambda$;

$E=hc/\lambda$, where c is the velocity of the radiation in a vacuum and λ is the wavelength.

Hence the larger the frequency [the shorter the wavelength] the greater the energy of a radiation.

Note that all the energy of one photon is absorbed by one electron. This implies that the energy of the radiation must be sufficient to dislodge an electron from the surface of the metal otherwise no electron would be emitted. Electrons of various metals require different amounts of energy to be emitted.

The minimum energy requirement of any metal to emit an electron is referred to as the **workfunction, w_0** of that metal. This implies that the radiation being used must meet a certain minimum frequency below which no photoemission occurs. This minimum frequency is called the **threshold frequency, f_0**.

Hence workfunction, $w_0=hf_0$.

For any radiation of frequency f which is less than the threshold frequency f_0 of the metal surface, the energy, hf of the radiation will be less than the workfunction, w_0 of the metal. Hence no photoemission takes place. However, when the frequency, f of the radiation is greater than the threshold frequency f_0 of the metal, then the amount of energy equilavent to the workfunction of the metal will be used to emit an electron and the rest of the energy will be converted into kinetic energy of the electron.

i.e energy of the radiation= workfunction + kinetic energy of the electron.

$E=w_0 + k.e$

$hf=hf_0+\frac{1}{2}mv^2$, where m- mass of an electron ($9.11\text{x}10^{-31}$kg) and v- the velocity of the electron.

This equation is known as Einstein's equation of photoelectric emission.

Alternatively, the radiation being used must not exceed a certain maximum wavelength for photoemission to occur [**recall $w_0=hf_0=hc/\lambda_0$**].this is called threshold wavelength.

The kinetic energy of an electron is sometimes expressed in terms of electron-volt (eV). It is the kinetic energy gained by an electron when it passes through a potential difference of one volt;

i.e**1eV=[1Vx($1.602\text{x}10^{-19}$C)]= $1.602\text{x}10^{-19}$J.**

Example 33.1

1. Calculate the energy of a photon of light of frequency $5.0\text{x}10^{14}$Hz in (a) joules (b) electron-volt.

   a) $E=hf=6.63\text{x}10^{-34}\text{x}5.0\text{x}10^{14}=3.315\text{x}10^{-19}$J

   b) {($3.315\text{x}10^{-19}$Jx1eV)/($1.602\text{x}10^{-19}$J)}= 2.0693eV

2. Light of frequency $4.3\text{x}10^{14}$Hz is irradiated on a surface of metal whose workfunction is 2.6eV. Explain whether photoelectric emission will occur or not.

   Energy of the radiation=hf=[$6.63\text{x}10^{-34}\text{x}4.3\text{x}10^{14}$]/ [$1.602\text{x}10^{-19}$J]= 1.7796eV

   Since the energy of the radiation is less than the workfunction, no photoelectric emission will occur.

   Alternatively;

   Workfunction=$hf_0=2.6\text{x}1.602\text{x}10^{-19}$

   $f_0=[2.6\text{x}1.602\text{x}10^{-19}]/[6.63\text{x}10^{-34}]= 6.2824\text{x}10^{14}$Hz

Since the frequency of the radiation is less than the threshold frequency of the metal, there is no photoelectric emission.

33.3: Factors affecting photoelectric emission

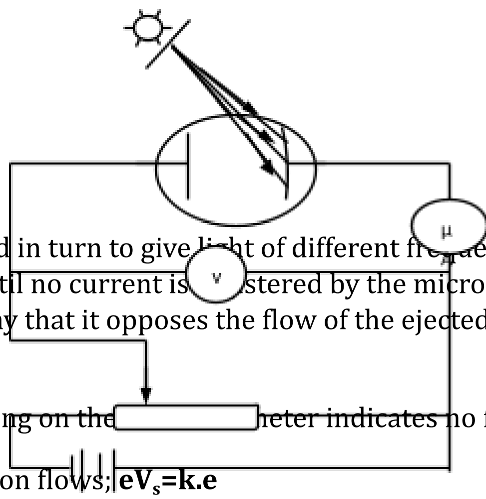
There are three main factors affecting photoelectric emission namely:

- Energy of the radiation
- Intensity of the radiation
- Type or nature of the metal

33.3.1: Energy of the radiation

The amount of energy of the emitted electrons is directly proportional to the frequency of the radiation. This can be shown by using radiations of different frequencies and investigating the stopping potential for each radiation. Stopping potential is the potential difference at which none of the emitted electrons reach the anode.

Colour filter



Various filters are used in turn to give light of different frequencies. For each filter, the variable resistor is used to vary the resistance until no current is registered by the micro-ammeter. Note that the source of d.c voltage is connected in such away that it opposes the flow of the ejected electrons i.e it works against the kinetic energy of the ejected electrons.

The absence of a reading on the micro-ammeter indicates no flow of electrons.

Hence, when no electron flows; $eV_s=k.e$

$eV_s=½mv^2$.

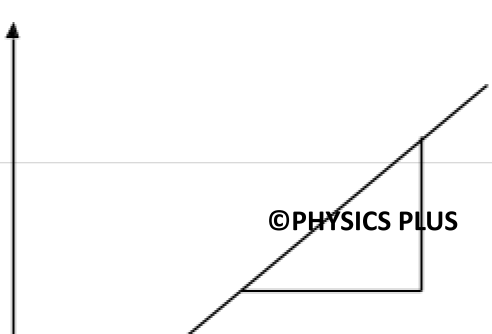
Substituted in the Einstein's equation, we obtain:

$hf=hf_0+eV_s$.

Below is a typical result obtained by using different colour filters of different frequencies and their corresponding stopping potentials:

Frequency, $f(x10^{14})$Hz	1.65	1.34	1.17	1.00
Stopping potential, V_s(V)	0.20	0.40	0.60	0.98

When a graph of the stopping potential V_s against frequency is plotted, the graph would appear as shown below:



Stopping potential **Slope=$\Delta V_s/\Delta f$ =h/e**

[V_s]

0

f_0 **Frequency, f[x10^{14} Hz]**

-w_0/e

33.3.2: Intensity of the radiation

It is defined as the rate of energy flow per unit area when the radiation is normal to the surface (area);

Intensity=E/At

But E/t =power, P.

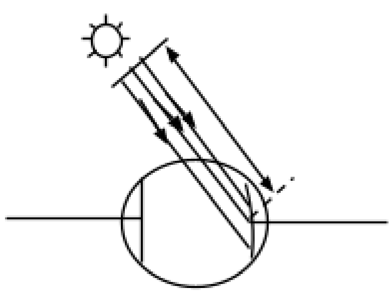
Hence, Intensity= P/t.

Suppose the source of the radiation is a distance **r** from the metal plate, then the intensity of the radiation is inversely proportional to the square of the distance **r**;

$I \alpha 1/r^2$.

Thus as the distance r decreases the intensity of the radiation increases and hence the value of the current is increased.

r



33.3.3: The type (nature) of the metal

Each metal has its own workfunction and hence threshold frequency. If the energy of the radiation striking the metal is below its workfunction then no electron will be ejected despite its intensity.

Example 33.2

1. The threshold wavelength of a photoemissive surface is 0.45μm. calculate:

a. The threshold frequency of the surface.

$f_0=c/\lambda_0=[3.0x10^8 m/s]/[0.45x10^{-6} m]$

$=6.67x10^{14}$Hz.

b. The workfunction of the surface in eV.

$W_0=hf_0=[6.63x10^{-34}x6.67x10^{14}]/[1.602x10^{-19}]=2.76eV$

c. The minimum speed with which a photoelectron is emitted if the frequency of the radiation is $7.5x10^{14}$Hz.

$6.63x10^{-34}x7.5x10^{14}=[6.63x10^{-34}x6.67x10^{14}]+[½x9.11x10^{-31}x v^2]$

$V=[12.081x10^{10}]^{½} =3.4754x10^5 m/s.$

9.4: Applications of photoelectric effect

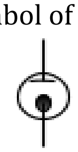
33.4.1: Photoemissive cell

It consists of a cathode and an anode. When light falls on a photosensitive cathode, electrons are dislodged which are then attracted by the anode. This completes an external circuit and current flows. When a body passes between the cathode and the source of light, the light is cut off and no photoemission takes place.

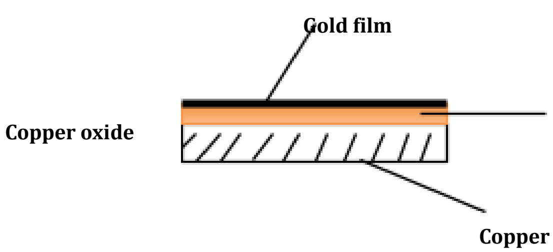
Such cells can be used in:

- ✔ Automatic opening of doors
- ✔ Burglar alarms for security
- ✔ Automated counting machine
- ✔ Reproduction of sound from a film.

Below is the symbol of a photoemissive cell.

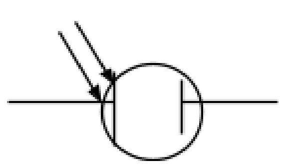


33.4.2: Photovoltaic cells



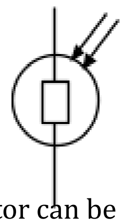
Light strikes the cell on the gold film side which emits electrons from the copper oxide surface. The copper oxide thus acquires a negative potential and copper a positive potential. A potential difference is therefore created and a current flows through a wire connecting the gold film and the copper externally.

Below is the symbol of a photovoltaic cell:



33.4.3: Photo-conductive cell

It is also called the **light dependent resistor.** The resistance of the cell varies with intensity of the light falling on it. In darkness, the resistance of the cell is greatest and least on a bright light. Below is the symbol of the photo-conductive cell:



Light dependent resistor can be used in operating street lights, fire alarms, detection and measurement of infra red radiation.

TOPIC 34: RADIOACTIVITY

34.1: Introduction

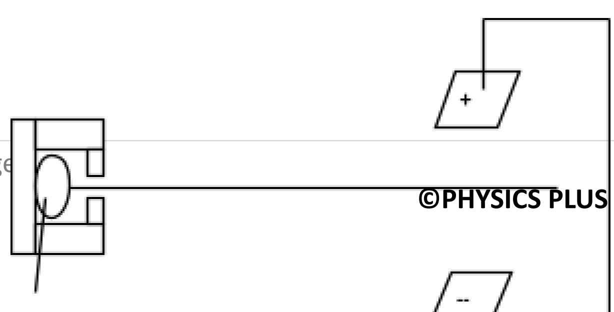
An atom X of mass number A and atomic number Z can be represented as. If the number of neutrons in the nucleus is N, then:

A= Z+N.

Some atoms have the same number of protons in the nucleus yet different mass numbers. Such atoms are referred to as isotopes. Examples of isotopes include carbon- 12 and carbon- 14. The energy holding the protons and neutrons together in the nucleus is called the **binding or nuclear energy.**

When the ratio of the number of protons to the number of neutrons in a nucleus is about 1:1, the nuclide is said to be stable, otherwise it is an unstable. For unstable nucleus, it has to undergo disintegration a process called radioactivity. Radioactivity is the spontaneous disintegration of the nucleus of unstable atom to release radiations.

In the process of radioactivity, there are three radiations which may be emitted namely alpha (α), beta (β) and gamma (γ) radiations. Their behavior can be observed when they are passed through a magnetic or electric field.



P

Q

R

Radioactive source

P- Beta radiation

Q- Gamma radiation

R- Alpha radiation

Alpha radiations:

- ✔ Are positively charged.
- ✔ Are massive or heavy and thus have shorter range in air. They are slightly deflected by strong magnetic or electric field due to their higher mass.
- ✔ Cause the highest ionization effect on the particles on their paths compared to beta and gamma radiations, thereby losing most of their energy.
- ✔ Have the least penetrating ability or power compared to the other two radiations. They can be stooped by a thick sheet of paper.

Beta radiations:

- Are negatively charged.
- Are lighter compared to alpha radiations. Hence they are greatly deflected by strong magnetic or electric field.
- Have longer range in air.
- Cause less ionization compared to alpha radiations. Hence they have a higher penetrating ability or power. They can penetrate a thick sheet of paper but can be stopped by a thin aluminium foil.

Gamma radiations:

- Are massless and do not have charge. Hence they are not deflected by both magnetic and electric fields.
- Are electromagnetic waves.
- Cause very little ionization. Hence most of their energy is intact. They have the highest penetrating ability or power of all the three radiations. They can penetrate thick paper and aluminium but is stopped by thick lead.

34.2: Radioactive decay and the decay equations

The original atom before the decay process is referred to as the **parent/mother** nuclide and the product is referred to as the daughter nuclide. A radioactive decay process consists of a parent nuclide, a daughter nuclide and the emitted radiation(s).

+ Emitted radiation(s)

Where X- the parent nuclide

Y- the daughter nuclide

Note that a particular radioactive decay process must not necessarily emit all the three radiations.

Suppose a radioactive decay process takes the form shown by the equation below:

+ , wh~~ere X is the p~~arent nuclide, Y is the daughter nuclide and Q is the emitted radiation;

Then, **A=m + a** and**Z=n + b**.

Radioactive decay is not dependent on physical factors like pressure, temperature or chemical composition of the nuclide.

There are three types of radioactive decay:

34.2.1: Alpha decay

This decay process emits alpha radiation(s). Alpha radiation is the nucleus of a helium atom represented by. If a nuclide decays by releasing an alpha particle, the mass number of the parent nuclide is reduced by 4 while atomic number is reduced by 2;

$\alpha \xrightarrow{+}$

Example 34.1

$\alpha \xrightarrow{+}$

$\alpha \xrightarrow{+}$

34.2.2: Beta decay

When an atom undergoes beta decay, it emits a beta particle. A beta particle is a fast moving electron represented by**.** The mass number of such a nuclide remains the same while its atomic number increases by one (1).

β+. $\longrightarrow$

Example 34.2

β + $\longrightarrow$

β + $\longrightarrow$

34.2.3: Gamma decay

Gamma decay does not have any effect on the mass number or atomic number of the nuclide. Instead the nuclide attains stability by simply releasing energy in the form of gamma radiation.

γ $\longrightarrow$ **+ gamma radiation**.

Example 34.3

$\gamma \xrightarrow{+\gamma}$

Example 34.4

1. Uranium, undergoes a decay emitting alpha and beta particles to become lead. Calculate the number of alpha and beta particles emitted.

$+ \text{ x } () + \text{y}() \longrightarrow$

228= 208+4x+0

x=20/4 =5 alpha particles

Also 92=82+2x-y

y=82+10-92=0

There was no beta particle emitted.

2. Uranium, decays to polonium, by emitting alpha particles. Write down the nuclide equation to represent the decay process. Hence determine the number of alpha particles emitted.

$\xrightarrow{\alpha} \; + \text{ x}()$

234= 218+4x

x=16/4 =4 alpha particles.

Alternatively, 92=84+2x

x=8/2 =4 alpha particles.

34.3: Radiation detectors

Below are some of the radiation detectors:

34.3.1: A photographic film

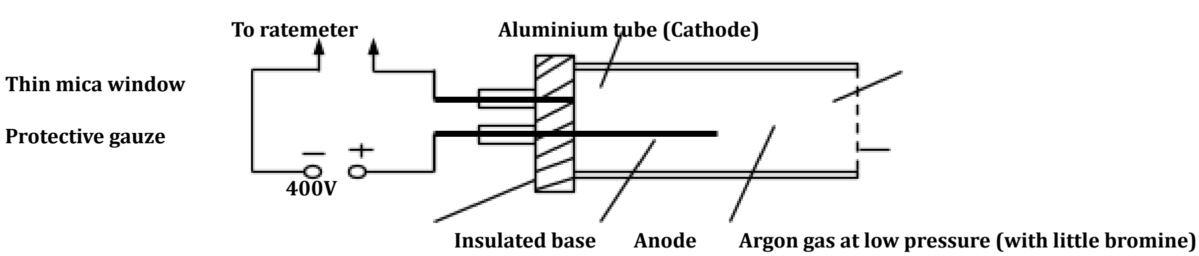
When radioactive radiations strike a photographic film, they cause photographic emulsion i.e the film is blackened.

34.3.2: The leaf electroscope

We have already seen that alpha and beta particles can ionize particles on their paths. This produces ions. If a source of these radiations is brought near the cap of a charged electroscope, the electroscope repels ions of similar charge but attracts those of the opposite charge. This neutralizes the electroscope and the leaf falls.

This method is most suitable for alpha particles since they cause the highest ionization but is not suitable for gamma radiations because they cause the least ionization.

34.3.3: The Geiger Muller (GM) tube



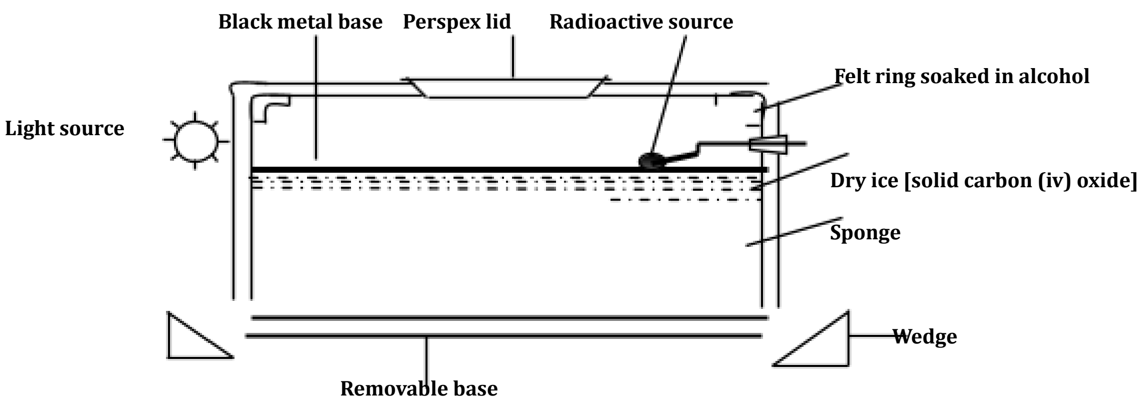
The radiation enters the tube through the thin mica window. The radiation ionizes argon gas. Opposite ions are attracted to either the cathode or the anode making a pulse of current to flow. As these ions move towards either electrode, they continue ionizing the argon gas producing more ions. The current is passed through an amplifier and then to a ratemeter where it is registered.

Note that only one pulse should be registered for each ionizing particle entering the tube. However, due to the high energy content of the positive ions, more electrons may be liberated from the surface of the cathode when struck by the positive ions. Such electrons are called **secondary electrons**. These can cause further ionization rendering the pulse registered incorrect.

To counter this, bromine is used which acts as a quenching agent, absorbing the energy of the positive ions before they reach the cathode.

This method is not suitable for detection of gamma radiations due to its low ionization effect.

34.3.4: The diffusion cloud chamber



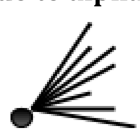
This detector uses the concept that when an ionizing radiation passes through air with saturated vapour, then the vapour is observed to condense on the ions formed. This explains why aeroplanes sometime leave trails of cloud behind them as they move through super saturated air.

In the diffusion cloud chamber, alcohol vaporizes and diffuses towards black metal base. When a charged particle from the radioactive source; either alpha or beta particle, knocks the air particles ions are produced. The vaporized alcohol condenses on the formed ions. Since positive ions are heavy, they remain behind forming tracks which can be clearly seen through the Perspex lid. To enhance visibility, a source of light is used to illuminate the chamber.

The dry ice is used to keep the black metal base cool while the sponge is used to keep the dry ice in contact with the black metal base.

Each radiation will produce a specific track as shown below:

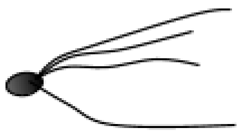
Tracks due to alpha radiation



They are:

- Short, indicating their shorter range in air.
- Straight; due to their mass it is not easy to displace them from their path by air particles.
- Thick, to show they are heavy particles.

Tracks due to beta radiation



They are:

- Long, indicating their longer range in air.
- Thin, indication of their lower mass.
- Irregular in direction (not straight), meaning that they can be displaced by air particles.

Tracks due to gamma radiation

Tracks due to gamma radiation are generally scanty and disjointed. These tracks do not come directly from the source but from electrons released by the gas atoms when they are struck by gamma radiation. The electrons then produce their own tracks.

34.4: Background radiation

Sometimes even in the absence of a radioactive source nearby, a GM tube may still register some radiations. This is called background radiation and it is present within the atmosphere. Some of the causes of background radiation include radioactive substances in air, ground and bricks of buildings, cosmic rays, sun's radiations, some rocks, natural and artificial radioisotopes etc.

34.5: The decay law

A radioactive decay occurs by chance i.e it is non-predictable. The decay states: **the rate of disintegration at any given time is directly proportional to the number of nuclides remaining undecayed**;

$\delta N/\delta t \ \alpha -N$

$\delta N/\delta t = -\lambda N$, where N-is the number of nuclides undecayed (remaining) and λ- is the decay constant.

Note that the negative sign indicates that the number N is decreasing with time.

$\delta N/\delta t$ is referred to as the **activity** of the material.

The above equation can be rearranged as;

$\delta N/N = -\lambda \delta t$

Suppose N_0 nuclides reduce to N nuclides between a time t=0 and t=T in a decay process, then by integration we have;

$\delta N/N = -\lambda \delta t$

$= -\lambda$

In N-In $N_0 = -\lambda[0-T]$

In $(N/N_0) = -\lambda[0-T] = \lambda T$

$N/N_0 = e^{\lambda T}$

$\mathbf{N = N_0 e^{\lambda t}}$

34.6: Half life, $t_{½}$

This is the time taken for half the number of nuclides initially present in a given radioactive sample to decay.

From the equation, $N = N_0 e^{\lambda t}$;

When $t = t_{½}$, $N = ½N_0$.

Then, $½N_0 = N_0 e^{\lambda t½}$

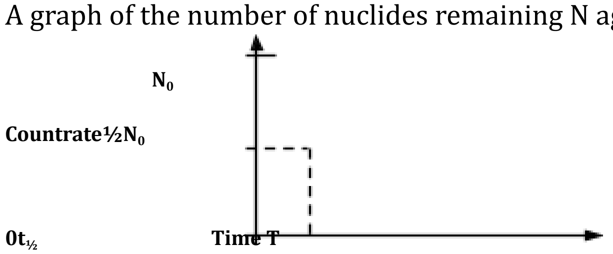
Thus In $½ = \lambda t_{½}$

$0.693 = \lambda t_{½}$

And $t_{½} = 0.693/\lambda$, where λ is the decay constant.

It can also be shown that, $\mathbf{N = N_0(½)^{T/t½}}$

A graph of the number of nuclides remaining N against time T appears as shown below:



In order to plot the correct graph, it is advisable to first subtract the background radiation if does exist from each countrate before plotting the values. This will ensure that only countrate due to the radioactive material is used to plot the graph. This is because the value of the background radiation usually fluctuates.

Example 34.5

1. A radioactive substance is found to have an activity of 360 counts per second. 30minutes later, it was 45counts per second. Determine its half life.

   $$\mathbf{360 \xrightarrow{t_{½}} 180 \xrightarrow{t_{½}} 90 \xrightarrow{t_{½}} 45}$$

   Hence $3t_{½}$=30minutes

   $t_{½}$=30/3 =10minutes.

Alternatively

$N=N_0(½)^{T/t½}$

$45=360(½)^{30/t½}$

$2^{-3}=2^{-30/t½}$

$-3=-30/t^{½}$

$t_{½}=-30/-3$ =10minutes

2. A radioactive substance has a half life of 10hours. Calculate the percentage of the sample that remains after 25hours.

   $N=N_0(½)^{25/10}$

   But percentage of the sample remaining after 25hrs is given by; [N/N0]*100

   Hence $[\{N_0(½)^{25/10}\}/N_0]*100$ =17.68%

3. A GM tube is used to measure the decay of a certain radioactive substance and the results are as shown in the table below. The background radiation is 25counts per hour.

Time (hrs)	**0**	**1**	**2**	**3**	**4**	**5**
Countrate (counts/h)	**425**	**255**	**175**	**105**	**73**	**51**

Plot a graph of countrate against time and use it to determine the half life of the material.

34.7: Nuclear fusion and fission

Nuclear fusion is where light nuclei combine to form a heavier nucleus. The process is accompanied by the release of large amounts of energy. Example is the fusion of lithium and hydrogen to give helium.

+ + + Energy $\longrightarrow$

Nuclear fission occurs when a nucleus splits into smaller more stable nuclei. This happens by the nucleus absorbing a neutron. During nuclear fission, the binding energy is released. Example is the fission of uranium-235;

+ + + 3() + Energy $\longrightarrow$ $\longrightarrow$

Nuclear fission is the principle on which hydrogen bombs work. This process if not controlled may lead to explosions.

34.8: Applications of radioactivity

In medicine:

- ✔ Gamma rays can be used to control cancerous growths in the human body.
- ✔ Gamma rays can be used to sterilize surgical equipment.
- ✔ Can be used to monitor blood circulation disorders and the functioning of thyroid gland.

In carbon dating- it uses the ratio of carbon-12 to carbon-14 to estimate the ages of fossils.

Pipe leakages- the content being transported through the pipe is mixed with some radioactive substance which can be detected by a radiation detector on the ground around the area of leakage.

In Agriculture- a radiation detector can be used to monitor the uptake of minerals introduced to plants by mixing it with some weak radioactive substance. Gamma rays can also be used to kill pests or make them sterile.

Determination of thicknesses of thin metal sheets, paper or plastics- a GM tube is used to measure the thickness of the metal plates, paper or plastic. The source of radiation is placed on one side while the GM tube is placed on the opposite side. The metal plate is passed between the source and the detector. The countrate registered is a measure of the thickness of the metal plate. To be more efficient, a thickness gauge can be adapted which automatically controls the thickness of the metal foils, paper or plastics.

34.9: Hazards of radioactivity and their remedy

The effects of radiation on a human body depends on:

- The nature of the radiation,
- Dosage and
- Part of the body irradiated.

Excessive exposure of body cells to radiations can lead to burn effects or genetic damage. Extreme heavy doses can be fatal. There could also be delayed effects such as cancer, leukemia and hereditary defects.

Gamma rays and beta radiation are more dangerous compared to alpha radiation due to their high degree of penetration.

Precautions should therefore be taken when handling radioactive materials. These include:

- Always use forceps to handle radioactive materials. Never use bare hands to hold such materials.
- Keep radioactive materials in thick lead boxes.
- Use radiation absorbers in hospitals and research laboratories.
- Reduce time spent near radiation sources.

TOPIC 35: ELECTRONICS

35.1: Introduction

This topic is about electronic circuits and their applications. Precisely it looks at the electrical conductivity of materials. Under this title, materials can be classified into three groups:

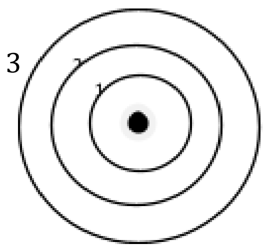
- Conductors
- Semiconductors
- Insulators

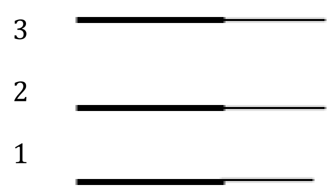
Conductors are those materials which allow current to flow through them easily. They are mainly metals like copper.

Materials which do not allow current to flow through them completely are referred to as **insulators**. They include plastic, paper, dry wood, rubber etc.

Semiconductors are those materials whose electrical conductivity lies between that of good conductors and insulators. They include silicon, germanium etc.

According to the **energy band theory**, when two atoms are brought closer to one another, the energy levels around the nucleus split into smaller sub-levels called **bands**. The outer energy level splits more easily giving many bands compared to the inner levels.





There are two important bands in any material which determine its electrical, optical and magnetic properties. These are the valence and conduction bands which are separated by a forbidden gap.

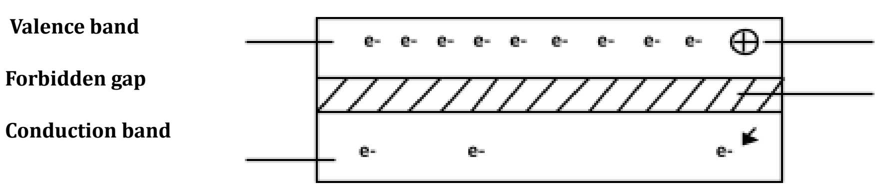
35.2: Semiconductors

Semiconductors are those materials whose electrical conductivity lies between that of good conductors and insulators. They have a smaller forbidden gap. The valence band is almost completely filled with electrons while the conduction band has almost no electrons.

When the temperature of the semiconductor is raised the electrons gain more energy and are able to move from the valence band across the forbidden gap to the conduction band. This increases the electrical conductivity of the semiconductor. Hence conductivity of semiconductors increases with temperature.

When an electron jumps from the valence band into the conduction band, a hole is left in its place. This is treated as a **positive charge**. Another electron in the valence band may jump into the hole formed creating another hole which may be filled by yet another electron and the process continues.

The movement of the electrons generates **electron current** while that of holes constitute **hole current**. Thus the net flow of current in semiconductors is due to the flow of electrons and holes.



There are two types of semiconductors as discussed below:

35.2.1: Intrinsic semiconductors

These are pure semiconductors whose electrical conductivity can be enhanced by increasing the temperature of the semiconductor. They include silicon, germanium etc. They have four electrons in their outermost energy level.

Their electrical conductivity is dependent on the electron-hole pair movement. The electrons and the holes are referred to as **charge carriers**. At room temperature, intrinsic semiconductors are insulators.

35.2.2: Extrinsic semiconductors

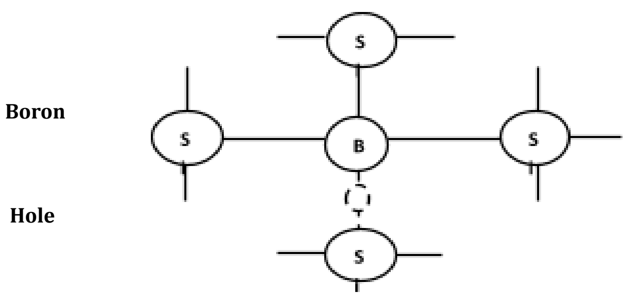
These are semiconductors obtained when a small amount of impurity is added to an intrinsic semiconductor. The process of adding an impurity to a pure semiconductor to improve its electrical conductivity is referred to as **doping**. Generally an extrinsic semiconductor is an impure semiconductor.

The impurity can either be a group three element e.g boron, gallium and indium or a group five element e.g phosphorous, antimony etc.

- **Doping using a group three element**

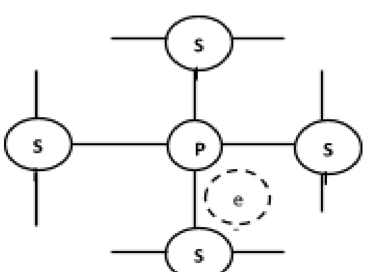
When silicon is doped using a group three element like boron, all the three electrons on the outermost energy level of boron atom participate in bonding with the neighboring atoms while silicon will have an extra electron. A vacancy will therefore exist due to the missing electron. This is treated as a **hole**. This hole is responsible for the electrical conductivity of the doped semiconductor. Hence holes are the majority charge carriers while electrons are the minority charge carriers. Such an impurity is called an acceptor impurity because they create a hole which can accept an electron.

An extrinsic semiconductor in which the majority charge carriers are holes is called a **p-type** semiconductor.



- **Doping using a group five element**

When a pure semiconductor is doped using a group five element like phosphorous having five electrons in their outermost energy levels, four of the electrons participate in bonding with the neighboring atoms while the remaining electron is used for electrical conductivity in the semiconductor. Hence electrons will be the majority charge carriers while holes will be the minority charge carriers. The impurity is referred to as a **donor impurity** since it donates an electron for electrical conductivity.

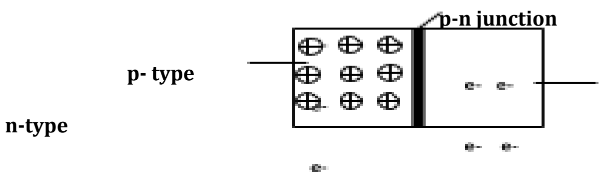


The resultant semiconductor is known as an **n-type** semiconductor.

Note that both p-type and n-type semiconductors are electrically neutral since the impurities added have the same number of electrons as there are protons.

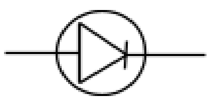
35.3: A P-N Junction diode

A p-n junction diode can be obtained when an intrinsic semiconductor is doped simultaneously using a trivalent and pentavalent impurities such that one half forms a p-type semiconductor while the remaining half forms an n-type semiconductor respectively. The boundary between the p-side and the n-side is referred to as a **p-n junction**.

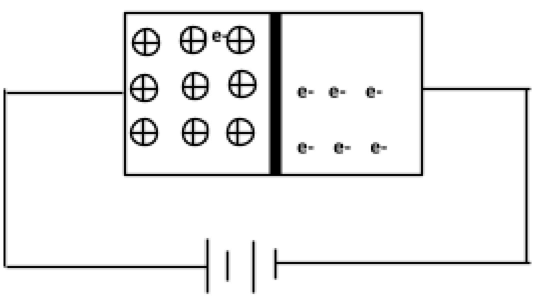


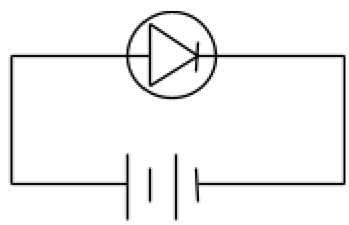
Immediately the junction is formed, a region called **depletion layer** is formed which prevents the free movement of electrons and holes across the junction. Thus the depletion layer develops a **potential barrier** at the junction. It acts as an insulator. For holes to cross to the n-side and electrons to the p-side, the potential barrier **must** be overcome.

The symbol of a p-n junction diode appears as shown below:



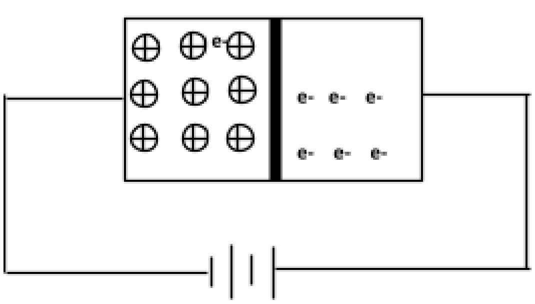
When a p-n junction diode is connected to a power supply it is said to have been **biased**. A p-n junction diode allows current to flow only in one direction when the p-side is connected to the positive terminal of the power source and n-side to the negative terminal of the power source. When connected this way, the diode is said to be **forward biased**.

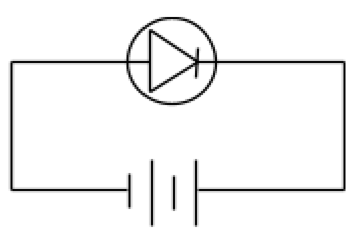




The cell provides the energy for the electrons to overcome the potential barrier and the holes are also able to cross over to the n-side thereby completing the circuit. The electrons and holes are attracted to the opposite ends. The thickness of the depletion layer is reduced and the charges flow with a lot of ease.

However, when the terminals of the cell are reversed such that the n-side is connected to the positive terminal and the p-side to the negative terminal of the cell, then the diode is said to be **reverse biased**.

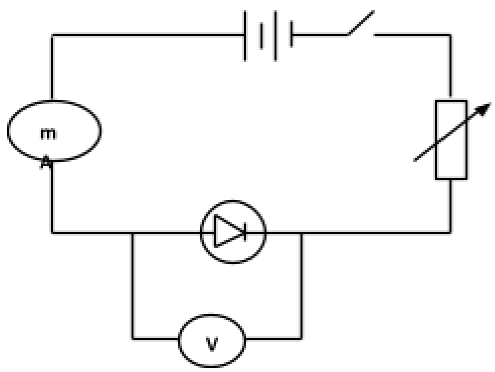




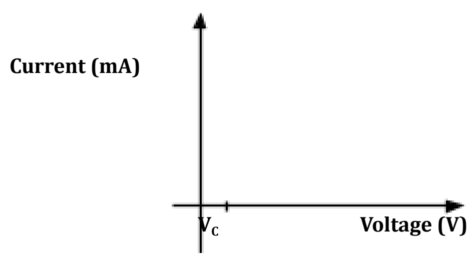
When the diode is connected in this manner, the holes in the p-type are attracted away from the junction by the external negative potential. Also, electrons are attracted away from the junction by the external positive potential. This increases the thickness of the depletion layer. Thus the potential barrier and hence the resistance of the function is increased. A very small current (leakage current) may flow in the circuit due to the flow of minority charge carriers.

35.4: Diode Characteristics

This is the relationship between current and voltage across a diode when connected to a power source. The set up below shows a circuit in which a diode has been forward biased:



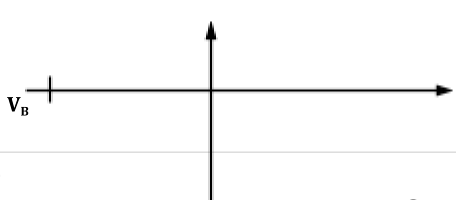
When the switch is closed, current flows through the diode since it is forward biased and it is recorded by the milliammeter. The voltage across the diode is measured by the voltmeter. The variable resistor is used to vary the current through the circuit. When a graph of current against voltage is plotted, the graph will be a curve as shown below:



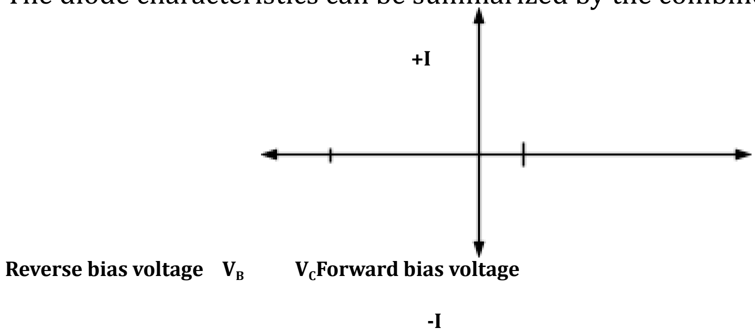
Initially as the forward voltage is increased from zero, no current is registered because the voltage is insufficient to overcome the potential barrier. When the potential barrier is completely overcome current start to increase. The voltage at which the potential barrier is overcome is referred to as the **cut-in voltage (V_c)**. Charges thereafter flow easily across the junction.

Since the graph is non-linear, it implies that a diode is non-ohmici.e it does not obey Ohm’s law.

However, when the diode is reverse biased a small current called leakage current flows. As the reverse voltage is increased the size of the current remains the same until a certain value when an appreciable amount of current starts to flow. This voltage is called the **zener/breakdown voltage (V_B)**. At this voltage the diode is damaged and therefore conducts electric current irrespective of the type of biasing.

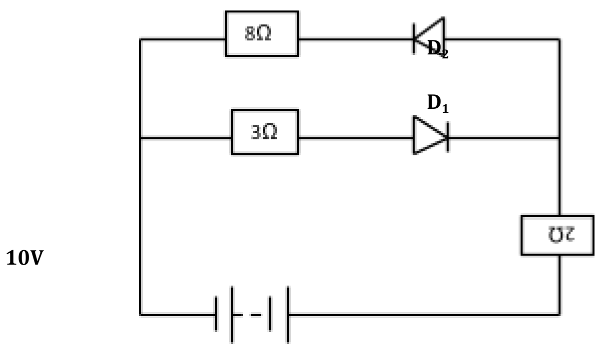


The diode characteristics can be summarized by the combined graph shown below:



Example 35.1

1. Calculate the current flowing through the 2Ω, 8Ω and 3Ω resistors in the figure below. Assume the diodes are ideal. Hence find the voltage drop across each resistor.



D_1 is reverse biased while D_2 forward biased.

Hence; $I_{3\Omega} = 0A$

$I_{2\Omega} = I_{8\Omega} = 10/(2+8) = 1A$

Also, $V_{3\Omega} = 0V$

$V_{2\Omega} = 1x2 = 2V$

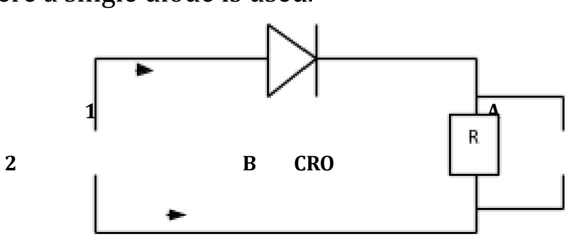
$V_{8\Omega} = 1x8 = 8V$

35.5: Applications of diodes

Diodes are used in rectification. Rectification refers to the process of conversion of alternating current to direct current. There are two types of rectification namely half-wave and full-wave rectification.

35.5.1: Half-wave rectification

Here a single diode is used.



During the first half cycle, the diode is forward biased. Current thus flows through the resistor R from the end A to B. During the second half cycle, the diode is reverse biased. No current flows through the diode and hence no current in the circuit. Hence in one complete cycle, one half is wasted. The output will appear as shown below:

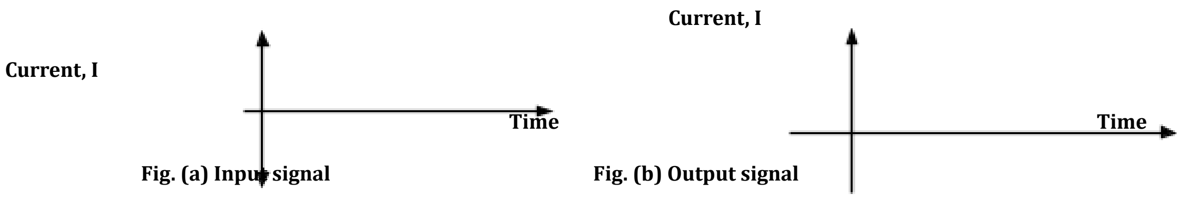


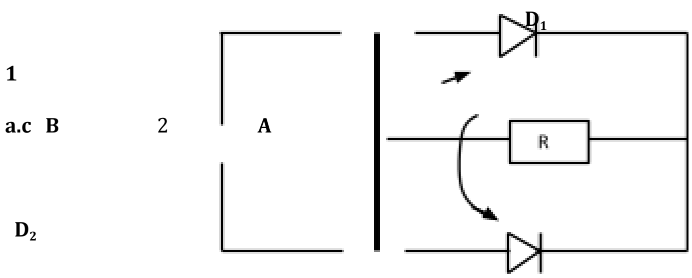
Fig. (a) Input signal   Fig. (b) Output signal

The output signal can be displayed on a CRO screen. In half-wave rectification, half of the input energy in every cycle is wasted. This limitation is eliminated in full-wave rectification.

35.5.2: Full-wave rectification

There are two ways of achieving full-wave rectification:

- **Using two diodes**



During the first half cycle, diode D_1 is forward biased while D_2 is reverse biased. Current thus flows through the resistor R from the end A towards B. During the second half cycle, diode D_2 will now be forward biased while D_1 reverse biased. Thus current flows through the resistor from the end A towards B.

Note that in both half cycles, the direction of flow of current through the resistor is the same. The resultant output will therefore take the form shown below:

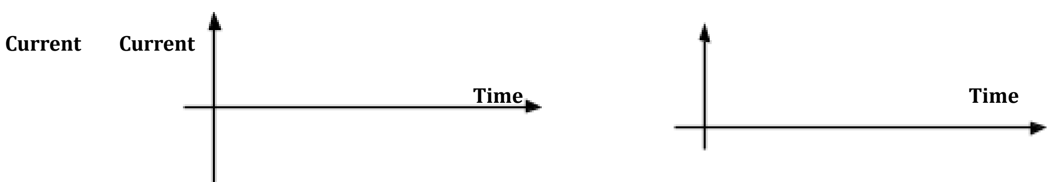
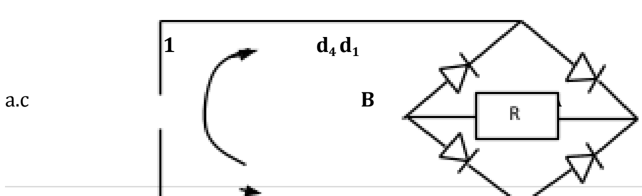


Fig. (a) Input   Fig. (b) Output

- **Using four diodes (bridge rectifier)**

A bridge rectifier uses four diodes such that in each half cycle two diodes are forward biased and the remaining two are reverse biased.

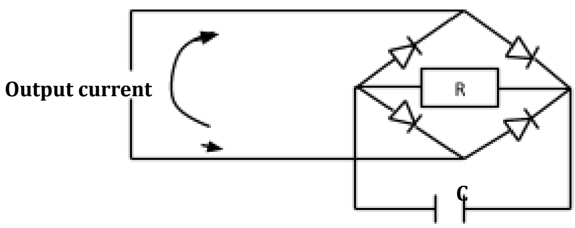


2 **d_3** **d_2**

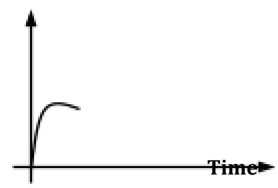
During the first half cycle, the diodes d_1 and d_3 are forward biased while d_2 and d_4 are reverse biased. Current thus flows through diode d_1 and d_3 via the resistor R. During the second half cycle, diodes d_1 and d_3 are now reverse biased while d_2 and d_4 are forward biased. Current thus flows through d_2 and d_4 via the resistor R.

Note that in both half cycles current flows through the resistor R in one direction only i.e from end A to B. This kind of rectifier can be used with very high voltages.

If a smooth rectified wave is needed, then a capacitor is connected across the resistor;



A smoothened wave



GOOD LUCK, NOW FACE THE KNEC