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PHYSICS FORM FOUR FULL NOTES

WAVES

INTRODUCTION

- A wave is a period disturbance which transfers energy from one place to another.
- There are two types of waves:

1. Mechanical waves
2. Electromagnetic waves

1. MECHANICAL WAVES

The mechanical waves are the waves which propagated through material medium such as solid, liquid or gas a speed which depends on the elastic and inertia properties of the material medium.

There are two types of mechanical wave;

(a)Longitudinal waves

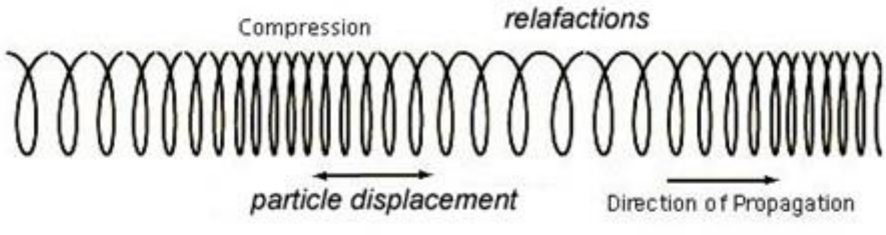
(b)Transverse waves

(a). LONGITUDINAL WAVES

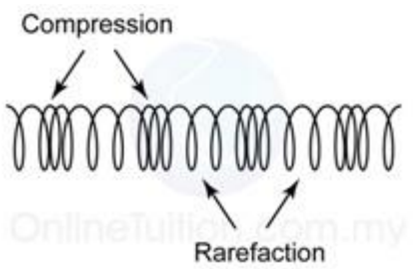
A longitudinal wave is a mechanical wave whose particle displacement is parallel to the direction of the wave's propagation .The particles in the medium are forced to oscillate along the same direction as that in which the waves is traveling.

Example;

If a horizontal loose stinky spring is set into vibration horizontally the waves travels horizontally.



NOTE: The regions along the material medium with the high pressure are called **compression** and the region with pressure is called **Rarefactions**.



(b). TRANSVERSE WAVES

- A transverse wave is a wave in which the direction of the wave's propagation is perpendicular to the direction of the particles displacement.

Example;

A loose horizontal slinky spring vibrates perpendicular to the waves which travels horizontally.

2. ELECTROMAGNETIC WAVES

Electromagnetic wave is a wave which does not necessary requires a material medium for its propagation and own also travel through vacuum. Examples of electromagnetic waves are radio waves, light waves, TV waves, x- ray, gamma rays, mobile phone waves etc.

The electromagnetic waves involves electric and magnetic field of the empty space vacuum acting perpendicular to each other

NB: The speed of all electromagnetic waves is **3.0×10^8m/s**

DIFFERENT BETWEEN MECHANICAL WAVES AND ELECTROMAGNETIC WAVES

MECHANICAL WAVES	ELECTROMAGNETIC WAVES
Can not be transmitted through a vacuum.	Can be transmitted even through vacuum.
They require material medium (solid liquid or gas) for propagation.	They do not require material medium for propagation.
Are causes by the vibrations of the particles of the material media through which they can pass.	Are caused by the effect of electric and magnetic field in the space.
Mechanical waves have low speed.	Electromagnetic waves have high speed.
Have long waves lengths.	Have short waves length.
Can be longitudinal or transverse waves.	Only transverse in nature.

WAVE PARAMETERS

A wave can be described fully by the following terms;

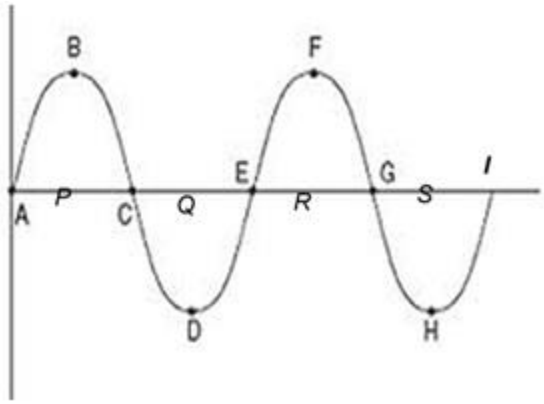
-Wave length

-Amplitude

-Time period

-Velocity

- Consider transverse waves **ABCDE** and **EFGHI** formed by a rope which one and is fixed to a pole and the other and is being moved up and down continuously.



1. **Wavelength**

- The distance between two nearest points on a waves which are in the same phase of vibration is called **wavelength** denoted by a Greek latter Lambda **(λ)** measured in **meters (m)**.
- Points **B** and **F** are crests, the distance between them is the wavelength.
- Points **D** and **H** are trough the distance between them is the wavelength.

NOTE

- A wavelength is the distance between two consecutive crests or trough of a wave.
- A wavelength is the horizontal distance completed by one cycle of waves.

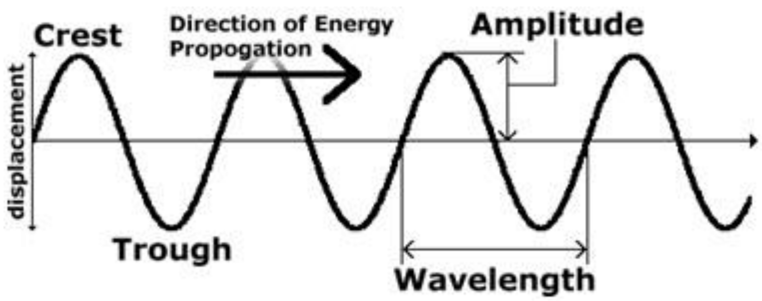
2. **Amplitude**

- An amplitude of a waves is a maximum displacement of particles of the material medium from their original undisturbed position.
- The amplitude of the wave denoted by the letter **A** measured in **maters (m)**.
- This quantity (amplitude) tells us about the size of the waves (big or small).

- The amplitude of a waves can be also defined as the height of the crest or depth of the trough (refer the diagram)

• The **BP**, **FR** are the amplitude of the waves the (the height of the crest).

• The **QD**, **SH** are the amplitudes of the waves (the depth of the through).



3. Time period

- Time period is the time neared to produce one complete waves or vibration or oscillation or cycle or to and fro motion.
- The time period of a waves is denoted by **T** measured in second **(s)**

4. **Frequency**

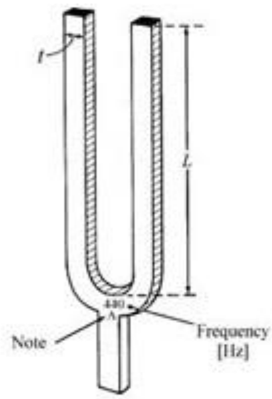
- The frequency of waves is the number of complete waves or vibration or oscillation or cycle produce in one second.
- It is denoted by **f** measured in **hertz** (**Hz**)
- If 5 complete cycles / waves / vibrations are produces in one second then the frequency is **5Hz**, if 100 complete vibrations are produces in one second then the frequency is **100Hz.**

NB: 100Hz mean there are 1000 complete waves being produced in one second.

Example:

- Tuning forks are often marked with numbers like 512 Hz, 384 Hz, 256Hz, etc.   These numbers signify the frequency of vibration oscillation or cycles or waves of the tuning forks. Tuning forks of 512

Hz will make 512 vibrations per second and emit 512 Hz complete sound waves per second. When it is hit on hard surface.



5. VELOCITY

- The velocity or speed of waves is the distance traveled or moved by waves in one second.
- The velocity or speed of a waves is denoted by **V** measured in meter per second **(m/s)**

THE RELATIONSHIP BETWEEN VELOCITY, TIME PERIOD, FREQUENCY, AND WAVES LENGTH

Recall;

$$\text{Velocity} = \frac{\text{Distance travelled}}{\text{Time taken}}$$

- If the distance traveled by a waves is numerically equal to its wavelength **(λ)**, Then the time taken by the wave is equivalents to time period **(T)**

$$\text{Thus; Velocity} = \frac{\text{wave length}}{\text{Time period}}$$

.

$$V = \frac{\lambda}{T}$$

BUT; $\text{frequency} = \frac{1}{T}$

From; $V = \lambda \times \frac{1}{T}$

From; $V = \lambda \times f$

Hence; $V = \lambda \times f$

$$\mathbf{V = \lambda f}$$

Where V = velocity

λ = wavelength

f = frequency

This equation is known as **WAVE EQUATION**

Example;

1. Calculate the velocity of the wave whose wavelength is 1.7×10^{-2}m and frequency 2×10^{14}Hz

Solution

Data given

$\lambda = 1.7 \times 10^{-2}$m

$f = 2\times10^{14}$Hz

From;

$V = \lambda f$

$= 1.7 \times 10^{-2} \times 2\ 10^{14}$

$= 3.4 \times 10^{12}$m/s

The velocity of the wave is 3.4 1×10^{12}m/s

2. Find the wavelength of sound wave whose frequency is 550Hz and speed is 330m/s

Solution

Data given

f = 550Hz

V = 330m/s

From

$V = \lambda f$

$$\lambda = \frac{v}{f}$$

$$= \frac{330}{550} = 0.6$$

The wavelength is 0. 6m

NB: The higher the frequency of a wave, the shorter the wavelength and the lower is the frequency on the wave, the longer is the wavelength.

3. The radio waves have a velocity of about 3.0×10^{8}m/s and the wavelength of 1500m. Calculate the frequency of these waves?

Solution

Data given

V = 3.0×10^8m/s

λ= 1500m

f = ?

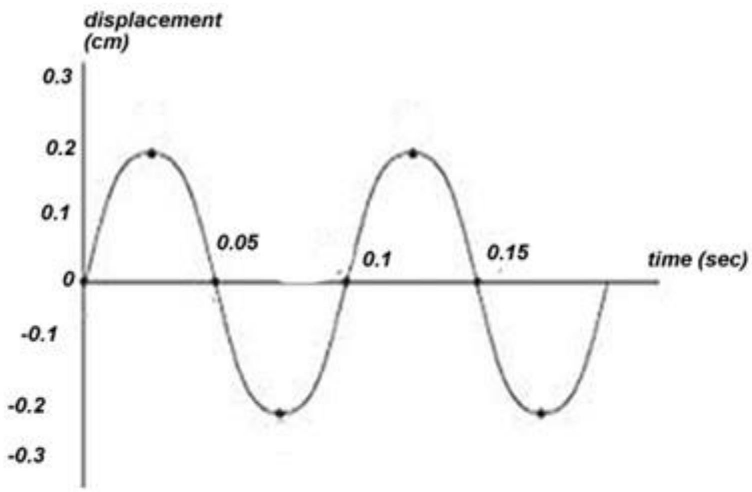
From: V = λf

$$f = \frac{v}{\lambda}$$

$$f = \frac{3.0 \times 10^8}{1500}$$

f = 2.0×10^5 Hz

4. The frequency is 2.0×10^5 Hz



The figure illustrates part of a wave traveling across the water at a particular place with velocity of 2m/s. Calculate;

1. The amplitude of the wave.
2. The frequency of the wave.
3. The wavelength of the wave.

(a) The amplitude of the wave is 0.2cm

b) $f = \frac{1}{T}$

1 complete cycle = 0.1 sec

$f = \frac{1}{0.1}$

$= 10Hz$

The frequency is 10Hz

c) $V = \lambda f$

$\lambda = \frac{v}{f}$

$= \frac{2}{10}$

The wavelength is 0. 2m

5. The wavelength of signals from a radio transmitter is 1500m and the frequency is the 200KHz. What speed to the radio wave travel?

•What is the wavelength of a transmitter operating at 1000KHz?

Solution

λ= 1500m

f = 200 KHz= 2, 00,000Hz

From;

V =λf

= 1500 x 200, 000

= 300, 000, 000 m/s

The velocity of the wave length is 3.0 x 10^8m/s

f= 1000KHz = 1,000,000 Hz = 1.0 x 10^6 Hz

V = 3.0 x10^8m/s

λ=?

V = λf

$$\lambda = \frac{V}{f}$$

$$= \frac{3.0 \text{ X } 10^8}{1.0 \text{ X } 10^6}$$

= 3. 0 x 10^2m

The wavelength is 3. 0 x 10^2m

6. A certain wave has time period of 0.04 second and travels at, 30 X 10^7 m/s Find its wavelength.

Solution

Data:

T = 0.04 sec

V = 30 x10^7m/s

=λ ?

From;

$$V = \frac{\lambda}{T}$$

λ = VT

= 30 x 10^7 x 0.04

= 3.0 x10^8 x 4.0 x10^{-2}

= 1.2 x10^7m

It is wavelength is 1.2 x10^7m

7. A personal with deep voice singing a note of frequency 200Hz is producing sound waves whose velocity is 330m/s.find the sound's wave length.

Solution

Data

f = 200Hz

V = 330m/s

From;

V=λ f

$$\lambda = \frac{V}{f}$$

$$\lambda = \frac{330}{200}$$

λ = 1.65m

8. The frequency of oxygen is 20 x 10^{13}Hz. find it's wavelengths.

Solution

Data

f = 20 x10^{13}Hz

V = 3.0 x 10^{8}m/s

From: V= λf

$$\lambda = \frac{V}{f}$$

$$= \frac{3.0 \times 10^{8}}{2.0 \times 10^{14}}$$

= 1.5 x 10^{-6}m

The wavelength is 1.5 x 10^{-6}m

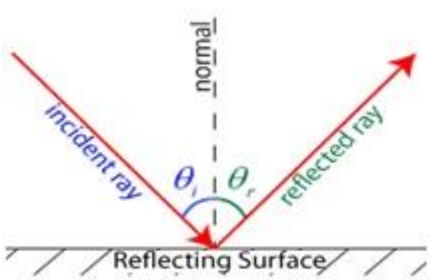
THE BEHAVIOR OF WAVES

All waves are general have similar properties (mechanical and electromagnetic). These behaviors include;

- Reflection of waves
- Refraction of waves
- Interference of waves
- Diffraction of waves

1. **THE REFLECTION OF WAVES**

- The Reflection of waves is the bouncing of waves or the sending back of waves on hitting the barriers.

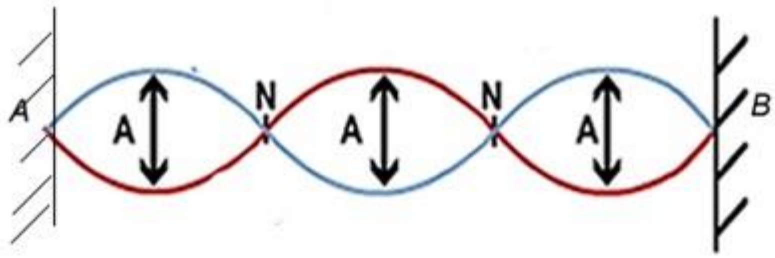


NB: The bouncing back of waves (example Sound waves) when striking hard surface, results onto the reflected sound known as **echo.**

An echo sound is the repetition of sound by the reflection of sound waves. An echo occurs when shouting in forests, in fronts of the tall buildings. In front of an escarpment, in front of mountain and in front of obstacles .

When an echo sound joins up with the original sound which then seems to be prolonged is known as **reverberation** (the multiple reflection of sound wave) this phenomena occur in large halls such as concert halls mosque, temples , cathedrals . In order to remove the reverberation in such hall they have to be equipped with **acoustic material** e.g. Papered walls, blankets, carpet, curtain, clothes, sponge material or any other material which can absorb the sound waves easily.

NOTE: The reflection of waves can be produce in strings. The standing waves or stationary waves formed when two or more traveling or progressive waves of the same frequency and amplitude travel in opposite direction.



N = Nodes

A = Anti nodes

- A standing wave or stationary wave is formed when an incident wave meet its reflection when in the medium. On the wave there are nodes and anti nodes.
- **A node** is point on a stationary wave which is completely at rest.
- **Anti nodes** is a point on a stationary wave which has a maximum displacement.
- The distance between two successive nodes;

$NN = \frac{\lambda}{2}$ (half a wave length)

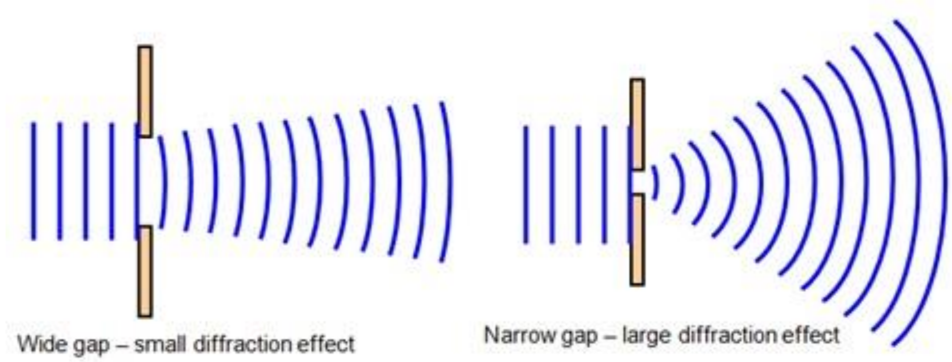
Also the distance between two successive anti nodes;

$$AA = \frac{\lambda}{2}$$

- The standing waves is formed by the process of **interference**
- **Interference** is a pattern formed when two wave overlap or meet in a medium.

2. **DIFFRACTION**

- **Diffraction** is the spreading of waves when they pass through a narrow opening or a sharp edge.
- A clearly diffraction is observed when the opening is about the size of the wave length of wave.
- In diffraction light spreads into the geometrical shadow.
- Sound can be heard round corners due to diffraction and the wavelength is comparable to the openings.

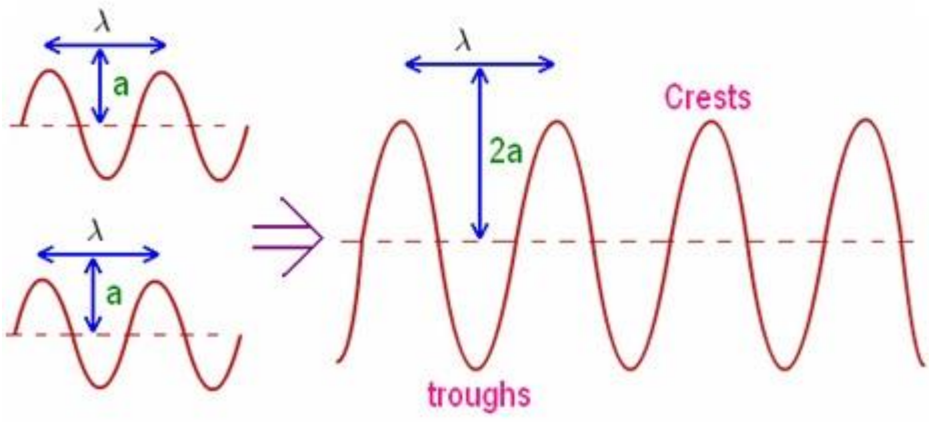


3. **INTERFERENCE**

- **Interference** is the pattern formed when two or more waves overlap in medium.
- Two waves overlap when they meet in a medium. They also superpose to superpose on each other.
- The pattern formed by interference is called an **Interference pattern**

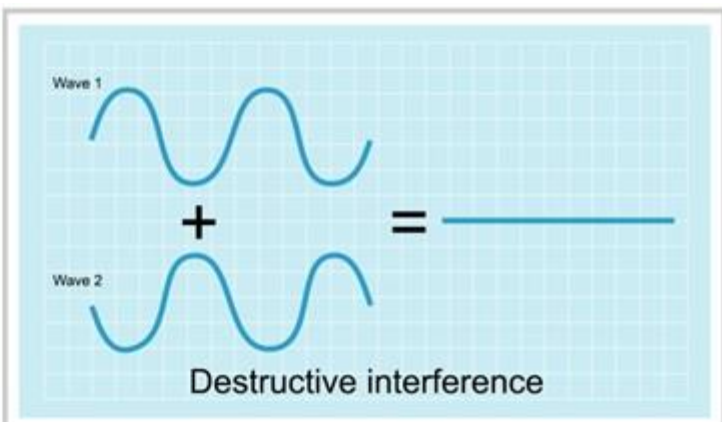
CONSTRUCTIVE AND DESTRUCTIVE INTERFERENCE

- **Constructive interference** occurs when a trough meets a trough or a crest meet crest producing maximum amplitude.



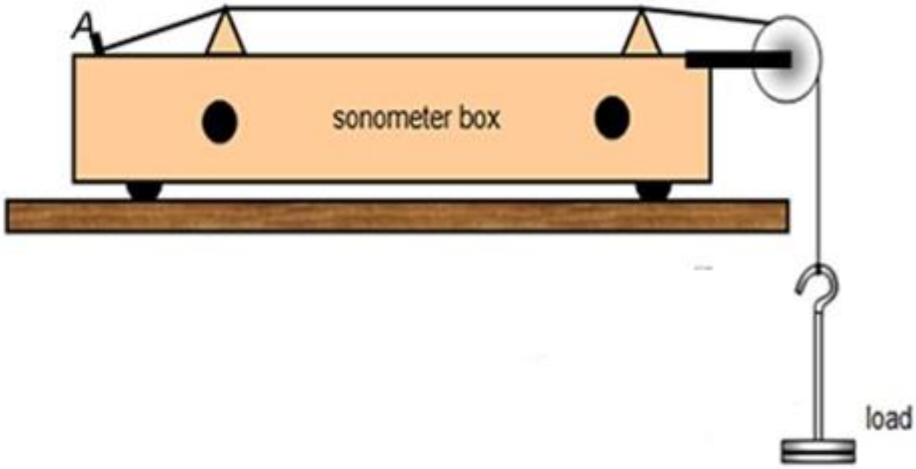
In this case the two waves are vibrating in the same phase/ direction

- **Destructive interference** occurs when a trough meet a crest. The resulting amplitude is smaller than the amplitude of the waves.
- The waves have equal amplitude , they cancel each other.



THE SONOMETER EXPERIMENT

- The sonometer is a device used to study the frequency and the velocity of the wave obtained from string instruments.



- The experiment have shown that the velocity of the wave produced from a string is directly proportional to the square root of the tension T of the string.

i.e. $V \propto \sqrt{T}$ ... ... (i)

- The velocity of a waves produce from a string instrument is directly proportional of the squares root of the length of the string

i.e. $V \propto \sqrt{l}$ ... ... (ii)

- The velocity of wave produce from a string instrument is inversely proportional to the squares roots of the mass of the string

i.e. $v \propto \sqrt{\frac{1}{m}}$ ... . (iii)

- The combination of expression (i) , (ii), and (iii) gives;

$$v \propto \sqrt{\frac{Tl}{m}}$$

$$v = K\sqrt{\frac{l}{m}}$$

But K = 1

So $v = \sqrt{\frac{Tl}{m}}$

If $\frac{m}{l}$ = linear density denoted by ρ

Hence

$$v = \sqrt{\frac{T}{\rho}}$$

Example1;

A string has a length of 75cm and a mass 0f 8.2g. The tension in the string is 18N. Calculate the velocity of the sound wave in the string.

Solution

Data given

Length L = 75cm = 0.75m

Mass M = 8.2g = 8.2×10^{-3}kg=

Tension T = 18N

$$from\ v = \sqrt{\frac{TI}{m}}$$

$$V = \sqrt{\frac{18 \times 0.75}{8.2 \times 10^{-3}}}$$

$$Recall\ \ f = ma$$

$$N = Kg/s^2$$

$$V = \sqrt{\frac{18\,Kg/s^2 \times 0.75}{8.2 \times 10^{-3} Kg}}$$

$$V = \sqrt{\frac{18 \times 0.75\, m^2/s^2}{8.2 \times 10^{-3}}}$$

$$V = \sqrt{\frac{13.5\, m^2/s^2}{9.2 \times 10^{-3}}}$$

$$V = \sqrt{1646\, m^2/s^2}$$

$$V = 40.5\, m/s$$

The velocity of the sound wave in the string was 40.5m/s.

2. Given that the velocity of the sound wave emitted from a string is 50m/s the Length of the string is 40cm and the mass of the string is 0.0004kg calculate the tension of the string.

Data given;

V = 50m/s

L = 40cm=0.4m

M = 4.0×10^{-4}kg

T=?

From;

$$V = \sqrt{\frac{TI}{m}}$$

$$50\,m/s = \sqrt{\frac{T\ \times 0.4}{4.0\ \times\ 10^{-4}}}$$

$$50 = \sqrt{\frac{0.4T}{4.0\ \times\ 10^{-4}}}$$

$$(50)^2 = \left(\sqrt{1.0\ \times\ 10^3 T}\right)^2$$

$$T = \frac{2500}{1.0 \times 10^3}$$

$$T = 2.5\ N$$

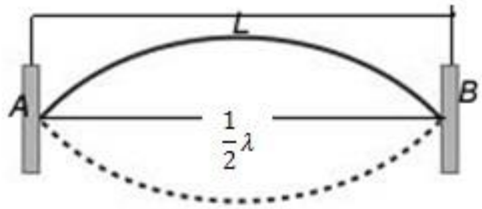
The tension of the string is 2.5N

NB: From the wave equation

$$V = \lambda f$$

$$f = \frac{v}{\lambda}$$

For the fundamental note (minimum frequency)



Solution

$L = \frac{1}{2}\lambda$

$\lambda = 2L$

HENCE

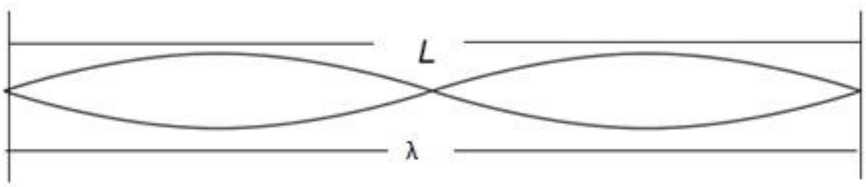
$f = \frac{V}{2L}$ (Minimum frequency)

Also

$f_0 = \frac{1}{2L}\sqrt{\frac{Tl}{m}}$

THE FIRST OVERTONE

Solution



If $L = \lambda$

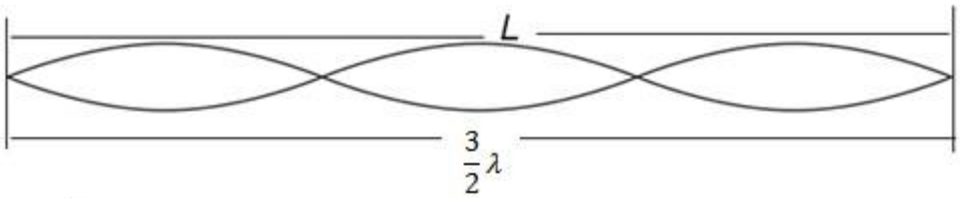
Then $f_1 = \frac{V}{L}$

$f_1 = \frac{1}{L}\sqrt{\frac{TL}{m}}$

Or $f_1 = \frac{1}{L}\sqrt{\frac{T}{\rho}}$

SECOND OVERTONE

Solution



If $L = \frac{3}{2}\lambda$

$f_2 = \frac{v}{\lambda}$

$f_2 = \frac{3v}{2L}$

HENCE

$f_2 = \frac{3}{2L}\sqrt{\frac{T}{\rho}}$

Generally:

$f_0 = \frac{1}{2L}\sqrt{\frac{Tl}{m}}$

$f_1 = \frac{1}{L}\sqrt{\frac{Tl}{m}}$

$f_2 = \frac{3}{2L}\sqrt{\frac{T}{\rho}}$

This can be formulated by the formula $\mathbf{f_n} = \frac{n+1}{2L}\sqrt{\frac{T}{\rho}}$

• Suppose that the tension on the liner density of the string are kept constant then

$f \propto \frac{1}{L}$

When two experiments conducted such that the lengths of the string are varied their corresponding frequency also varies.

Experiment1;

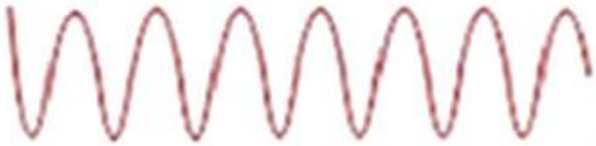
Lower frequency higher wavelength

$f_1 \propto \frac{1}{L_1}$

$f_1 = \frac{K}{L_1}$

$K = f_1 L_1 \ldots\ldots\ldots (i)$

Experiment2;



High frequency

Higher frequency smaller wavelength

$$f_2 \propto \frac{1}{L_2}$$

$$f_2 = \frac{K}{L_2}$$

Hence

$$f_1 L_1 = f_2 L_2$$

Also

$$\frac{f_1}{L_1} = \frac{f_2}{L_2}$$

Examples:

1. A sonometer wire of length 50cm vibrate with frequency 384Hz. Calculate the length of the sonometer wire so that it vibrates with frequency of 512Hz.

Solution

Data given:

L_1= 50cm

F_1 = 384Hz

F_2= 512Hz

L_2 = ?

From;

$$f_1L_1 = f_2L_2$$
$$384 \times 50 = 512 \times L_2$$
$$\frac{19200}{512} = \frac{512 \text{ x } L_2}{512}$$

L2=37.5cm

The length of the sonometer wire is 37.5m

2. A sonometer wire of length 40cm between two bridges produces a note of frequency 512Hz when plucked at midpoint. Calculate the length of the wire that would produce a note of frequency 256Hz with the some tension.

Data:

L_1 =40cm

f_1 =512Hz

L_2 =?

f_2 =256Hz

From;

$$f_1L_1 = f_2L_2$$
$$512 \text{ x} 40 = 256 \text{ x} L_2$$
$$\frac{20480}{256} = \frac{256 \text{ x } L_2}{256}$$

L_2 = 80cm

The length of the wire is 80cm

NB: Suppose that a frequency a wave produced from a string or wire is varied with tension (length and the linear density of the string or wire are kept constant).

$$f = \left(\frac{n+1}{2L}\right)\frac{\sqrt{T}}{\sqrt{\rho}}$$

$$f = k\sqrt{T}$$

So that $f \propto \sqrt{T}$

When two experiment is conducted to show the relationship between the frequency and the tension of the string.

Experiment 1: High tension , T

$$f_1 \propto \sqrt{T_1}$$

$$f_1 = k\sqrt{T_1}$$

$$K = \frac{F_1}{\sqrt{T_1}} \dots\dots\dots\dots (i)$$

Experiment2; low tension T_2

$$f_2 \propto \sqrt{T_2}$$

$$f_2 = k\sqrt{T_2}$$

$$K = \frac{F_2}{\sqrt{T_2}} \dots\dots\dots\dots (ii)$$

Hence:

$$\frac{f_1}{\sqrt{T_1}} = \frac{f_2}{\sqrt{T_2}}$$

ALSO;

$$f_1\sqrt{T_2} = f_2\sqrt{T_1}$$

Example1;

The frequency obtained from a plucked string is 400Hz when the tension is 2 Newton. Calculate;

a) The frequency when the tension is increased to 8N

b) The tension needs to produce a note of frequency 600HZ.

Data:

F_1 = 400Hz

T_1 = 2N

T_2 = 8N

From;

(a) $\frac{f_1}{\sqrt{T_1}} = \frac{f_2}{\sqrt{T_2}}$

$$\frac{400}{\sqrt{2}} = \frac{f_2}{\sqrt{8}}$$

$$f_2 = \frac{400 \times \sqrt{8}}{\sqrt{2}}$$

$$f_2 = 800\text{Hz}$$

The frequency is 800Hz

(b) $\frac{f_1}{\sqrt{T_1}} = \frac{F_2}{\sqrt{T_2}}$

$$\frac{400\text{Hz}}{\sqrt{2}} = \frac{600\text{Hz}}{\sqrt{T_2}}$$

$$400\sqrt{T_2} = 600\sqrt{2}$$

$$\sqrt{T_2} = \frac{600 \times \sqrt{2}}{400}$$

$$\left(\sqrt{T_2}\right)^2 = \left(\frac{3\sqrt{2}}{2}\right)^2$$

$$T_2 = \frac{9 \times 2}{4}$$

T_2 = 4.5N

The Tension is 4.5N

1. Given that the frequency obtained from a plucked string is 800Hz when the tension is 8N. Calculate;

   a) The frequency when the tension is doubled

   b) The tension required when the frequency is halved

Data:

T_2= 16N

T_1 = 8N

From

$$f_1\sqrt{T_2} = f_2\sqrt{T_1}$$

$$800\sqrt{16} = f_2\sqrt{8}$$

$$f_2 = 800 \times \sqrt{\frac{16}{8}}$$

$$f_2 = 800 \times \sqrt{2}$$

F_2 = 800 x 1.414

F_2 = 1131.2Hz

$$(b) f_1\sqrt{T_2} = f_2\sqrt{T_1}$$

$$800 \times \sqrt{T_2} = 400 \times \sqrt{8}$$

$$\sqrt{T_2} = \frac{1}{2} \times 2\sqrt{2}$$

$$\sqrt{T_2} = \sqrt{2}$$

$$T = \sqrt{2N}$$

CLASS WORK

1. Under constant tension the note produced by a plucked string is 300Hz when the length 0.9m;

a)At what length is the frequency 200Hz?

b)What frequency is produce at 0.3m

Data

F_1 = 300Hz

L_1 = 0.9m

a) F2= 200Hz
L2=

From;

$F_1 L_1 = F_2 L_2$

300 x 0.9 = 200 x L_2

$$L_2 = \frac{270}{200}$$

L_2 = 1.35m

b) L2 = 0.3m

$$f_1 L_1 = f_2 L_2$$

$$300 \times 0.9 = f_2 \times 0.3$$

$$f_2 = \frac{270}{3}$$

f = 90Hz

2. A string fixed between two supports that are 60cm a part. The speed of a transverse wave in a string is 420m /s. Calculate the wavelength and the frequency for;

i)Fundamental note

ii)Second overtone

iii)Fifth overtone

Data

L = 60cm

V = 420m/s

i) $f_n = \left(\frac{n+1}{2L}\right) v$

$f_0 = \left(\frac{0+1}{2\times0.6}\right) \times 420$

$f_0 = \frac{1}{1.2} \times 420$

$F_0 = 350$ Hz

From;

$V = \lambda f$

$\lambda = \frac{v}{f}$

$\lambda = \frac{420}{350}$

$\lambda = 1.2m$

The fundamental note is 350Hz and the wavelength is 1.2m

ii) $f_2 = \left(\frac{2+1}{2\times0.6}\right) \times 420$

$f_2 = \frac{3}{1.2} \times 420$

$F_2 = 1050$ Hz

From; $\lambda = \frac{v}{f}$

$\lambda = \frac{420}{1050}$

$\lambda = 0.4m$

The second overtone is 1050Hz and the wavelength is 0.
4m

iii) $f_5 = \left(\frac{5+1}{2\times0.6}\right) \times 420$

$f_5 = \frac{6}{1.2} \times 420$

$F_5 = 2100Hz$

From; $\lambda = \frac{v}{f}$

$\lambda = \frac{420}{2100}$

$\lambda = 0.2m$

The fifth overtone is 2100Hz and the wavelength is 0.2M

3. A string is fixed two ends 50cm a part. The velocity of a wave in a string is 600m/s. Calculate;
    1. The first five over tone
    2. The tenth five overtones

Data:

L = 50cm = 0.5m

V = 600m/s

(i) $f_n = \left(\frac{n+1}{2L}\right)V$

$f_1 = \left(\frac{1+1}{2\times 0.5}\right)X\ 600$

$f_1 = \frac{2}{1}\times 600$

$f_1 = 1200Hz$

(ii) $f_2 = \left(\frac{2+1}{2\times 0.5}\right)X\ 600$

$f_2 = \frac{3}{1}\times 600$

$f_2 = 1800Hz$

(iii) $f_3 = \left(\frac{3+1}{2\times 0.5}\right)^{600}$

$f_3 = \frac{4}{1}\times 600$

$f_3 = 2400Hz$

(iv) $f_4 = \left(\frac{4+1}{2\times 0.5}\right)^{600}$

$f_4 = \frac{5}{1}\times 600$

$f_4 = 3000Hz$

(v) $f_5 = \left(\frac{5+1}{2\times0.5}\right)^{600}$

$$f_5 = \frac{6}{1} \times 600$$

$$f_5 = 3600\text{Hz}$$

(2) The tenth and twelve overtones

$$f_{10} = \left(\frac{10+1}{2\times0.5}\right)^{600}$$

$$f_{10} = \frac{11}{1} \times 600$$

$$f_{10} = 6600\text{Hz}$$

$$f_{12} = \left(\frac{12+1}{2\times0.5}\right)^{600}$$

$$f_{12} = \frac{13}{1} \times 600$$

$$f_{12} = 7800\text{Hz}$$

• The first overtone is 1200Hz, 1800Hz, 2400Hz, 3000Hz, and 3600Hz.

• The tenth overtone is 6600Hz and the tenth overtone is 7800Hz.
NOTE: In stationary wave a string does note compose up to ten overtones, though mathematically is possible. In real practical of the sonometer by using turning, is possible for the second and third overtone.

CLASS ACTIVITY

1. Given that the refractive index of glass is 1.52. The wavelength of the radio waves in vacuum is 1.5 x 10^3m . Calculate the wavelength of the radio waves in glass.

$$\text{Refractive index} = \frac{\text{wavelenght in air}}{\text{wavelenght in glass}}$$

$$1.52 = \frac{1.5\times10^3}{W_G}$$

$$W_G = \frac{1.5\times10^3}{1.52}$$

$$W_G = 0.9868 \times 10^3$$

- The wavelength is 986.8m

1. A guitar wire fixed between two supports 60cm a part produces a wave of frequency 500Hz. Calculate;

(a)The frequency of a wave when the length of the guitar wire is reduced to quarter

(b)The length of the guitar wire when the frequency of the wave produced is 2000Hz

Data

L =60cm

F = 500Hz

(a)**From**

F1 = 500Hz

L_1 = 60cm

L_2 = 15cm

(a) From

$$f_1 L_1 = f_2 L_2$$

$$500 \times 60 = f_2 \times 15$$

$$\frac{30000}{15} = \frac{15 \times f_2}{15}$$

$$f_2 = \frac{30000}{15}$$

= 2000Hz

- The frequency is 2000Hz

1. F_1 = 500Hz

L_1 = 60cm

L_2 = ?

F_2 = 2000Hz

From;

$$f_1L_1 = f_2L_2$$

$$500 \times 60 = 2000L_2$$

$$\frac{50000}{2000} = \frac{2000 \times l_2}{2000}$$

L_2 = 15cm

- The length of the wire is 150m

Difference between sound wave and radio waves

Sound waves	**Radio waves**
• Are mechanical waves • Have low frequency • Have low velocity •Required material medium to prorogate	• Are electromagnetic waves • Have high frequency • Have high velocity •Do not required material medium to prorogate

Difference between longitude waves and transverse waves

Longitudinal waves	**Tran serves waves**
•Particle displacement is parallel to the direction of wave propagation	• Particle displacement is perpendicular to the direction of wave propagation

Describe briefly the phenomenon REVERBERATION

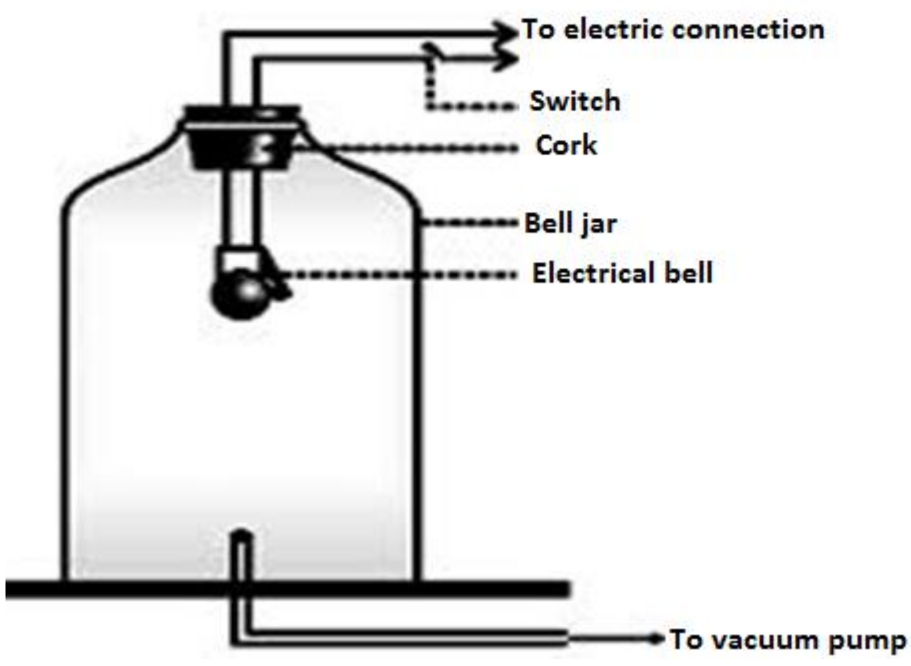
- When an echo sound Joins up with the original sound which then seems to be prolonged is known as **REVERBERATION**.

SOUND WAVES

- Sound waves are due to vibration of the particles of air or any other media in which they travel.
- An isolating body like a stretched string violin, drum, guitar, piano, vocal cords of human beings is disturbed to produce sound. The sound requires material media or matter in which they travel or propagate.

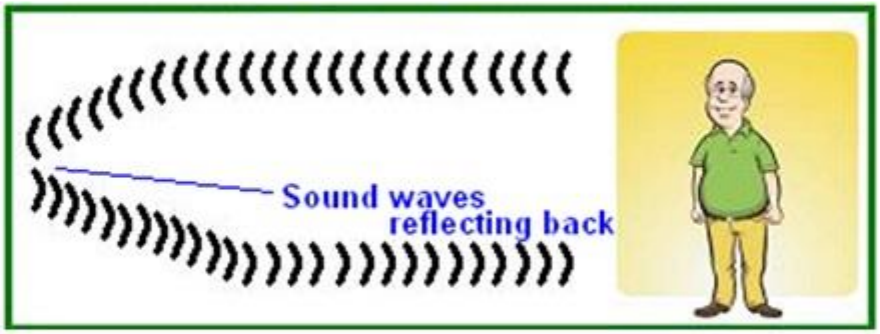
PROOF:

Sealing an electronic bell in a bell jar. Starting the bell ringing and then pumping out the air, the sound gravelly dies down. But the clapper can still be seen striking the gong. Allow the air to return the sound is heard again. This shows that sound needs medium of travel.



THE VELOCITY OF SOUND IN AIR

- Sound takes some time to travel from the sound producing body to our ears.
- Consider A to be a sound producing body and B is the tall vertical wall some meters away from A.



- When the sound is produced A it travels X meter to the wall and then back to point A covering the some distance XM in time sec. (To and from motion). The total distance moved by the sound wave in air is 2x (to and from motion) if velocity is given by;

$$\text{velocity} = \frac{\text{displacement}}{\text{total time}}$$

$$v = \frac{2d}{t}$$

Example;

1. Sound travelling towards a cliff 700m away takes 4.2 seconds for an echo to be heard. Calculate the velocity of sound in air.

Data

d = 700m

t = 4.2sec

From;

$$v = \frac{2d}{t}$$

$$v = \frac{2 \times 700}{4.2}$$

= 333.33m/s

- The speed of sound is 333.33m/s

2. A boy standing 100m from the foot of a high wall claps his hands and the echo reaches him 0.5 second later. Calculate the velocity of sound in air using this observation.

Data

d = 100m

t = 0. 5sec

From;

$$v = \frac{2d}{t}$$

$$v = \frac{2 \times 100}{0.5}$$

v = 400m/s

- The velocity of sound in air is 400m/s

3. A student standing between two vertical walls and 480m from the nearest wall, shouted. She heard the first echo after 3 seconds and the sound two second later use this information to calculate;

i) Velocity of sound in air

ii) Distance between the two walls.

Data

1. Velocity of sound in air

d = 480m, t =3 sec and v =?

From;

$$v = \frac{2d}{t}$$

$$v = \frac{2 \times 480}{3}$$

= 320m/s

- The velocity of sound in air is 320m/s

2. Distance between two walls

V = 320m/s, T = 5sec, d =?

$$v = \frac{2d}{t}$$

$$320 = \frac{2d}{5}$$

d =800m

Distance = $d_1 + d_2$

= 480 + 800

= 1280m

- The distance between the two wall is 1280m

4. An old woman sitting in a gorge between two large cliffs gives a short sharp sound. She hears two echo, the first after 1 second and the next after 1.5sec. The speed of sound is 340m/s what is the distance between the two cliffs?

Data

T_1 = 1sec

T_2 = 1.5sec

V = 340m/s

From;

$$v = \frac{2d_1}{t}$$

$$340 = \frac{2d_1}{1}$$

$$\frac{340}{2} = \frac{2d_1}{2}$$

$$d_1 = 170\text{cm}$$

From; $v = \frac{2d_2}{t_2}$

$$340 = \frac{2d_2}{1.5}$$

$$\frac{340 \text{x} 1.5}{2} = \frac{2d_2}{2}$$

d_2 = 255

Distance = $d_1 + d_2$

=170 + 255

=425m

- The distance between the cliff is 425m

5. A sonar signal (a high frequency sound wave) sent vertically downwards from the ship is refracted from the ocean floor and detected by a microphone on the keel. 0.4 sec after transmission. If the speed of sound in water is 1550m/s. What is the depth of the ocean in maters?

Data

T = 0.4sec

V = 1500m/s

From;

$v = \frac{2d}{t}$

$1500 = \frac{2d}{0.4}$

$2d = 1500 \text{ X } 0.4$

$d = \frac{1500 \times 0.4}{2}$

d = 300m

- The depth of the ocean is 300m

6. A man sees steam coming out from a factory whistle and 3 seconds later he hears the sound. The velocity of sound in air is 360m/s. Calculate the distance from the man to the factory.

Data

T = 3second

V = 360m/s

D = ?

From $v = \frac{d}{t}$

$360 = \frac{d}{3}$

d = 1080

- The distance from the man to the factory is 1080m

THE FACTOR AFFECTING THE VELOCITY OF SOUND WAVES IN DIFFERENT MATERIAL MEDIA

- The following are the factor influencing the velocity of sounds wave

1. The velocity of sound depends on the nature of material medium through which it travels. The speed of sound in air is about 340m/s. The speed of in water is about 1500m/s. The speed of sound in iron is about 5130m/s. Thus sound travels slowest in gases, faster in liquid and fasted in solids.
2. The velocity of sound depends on the temperature. As the temperature of the material media rises air. The speed of sounds at 0^0c is about 332m/s. At 20^0c the velocity of sound is about 340m/s. The speed of sound in air on a hot day is more than the speed of sound in a cold day.
3. The speed of sound depends on the humidity of air. The speed of sound is less in dry air. The speed of sound in air is more in humidity air as the humidity of air increases, the velocity of sound increases.

AUDIBILITY RANGE

- Human beings can hear sounds with frequency from about 20Hz to 20,000Hz (or 20Hz to 20 KHz). These are the limits for audibility, the upper limit decreases with age. The sound with frequency below 20Hz is known as infrasonic sound and the sound with the frequencies above 20 KHz are known as **ultrasonic sounds**.

- A bat can hear sounds with frequency above 20 KHz (**ultrasonic frequency**). Rats can hear sound with frequency below 20Hz (**infrasonic frequency**).

ECHO – LOCATION PRINCIPLE

- A bat has an acute vision in darkness. A bat emits the infrasonic sound from its mouth and noise which it holds open as it flies. It travels through air as a wave and the energy of this waves bounces off any object it comes a cross. A bat emits sound waves and listens very carefully to the echo. That returns to it. The bats brain processes the returning information by determining how long it takes the noise to return, the bats brain figures out how far away an object is. The bat can also determine where the object is how big it is and in what direction it is moving. The bat can tell if the object is to the right or left. By comparing the sound which reaches its right ear and left ear. It can easily turn according to avoid hitting the obstacles.

PROPERTIES OF MUSICAL SOUNDS

- The sound waves which produce pleasant sensation to our ears and are acceptable are called **musical sound**. The sound waves which produce troublesome sensations and are unacceptable are called **noise**. (Non – musical sounds).

- The sound waves which are produced by regular period vibration are musical in character while the sound waves which are produced by irregular non periodic vibration are non musical in character. There are three main characteristics by which one musical note is differentiated from other musical notes. These include:

- Pitch
- Loudness and intensity
- Quality of sound or timber of sound

1. **PITCH**

Pitch is the property of sound waves which helps us to differentiate between two sounds with equal loudness coming from different sources with different frequency. The pitch of sound depends on frequency. I.e. high frequency (high pitch) low frequency (low pitch).

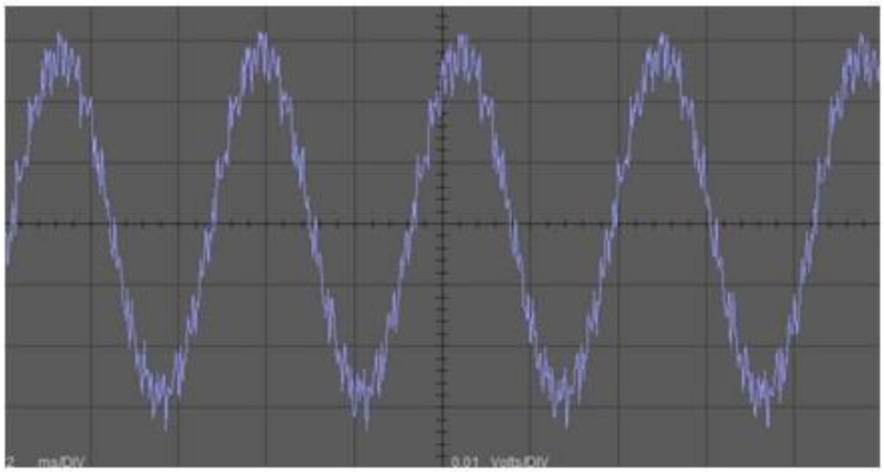
2.LOUDNESS OR INTENSITY

The loudness of the sound is the magnitude of the auditory sensation. The intensity sound is the time rate at which the sound energy follows through a unit area. The loudness depends on the amplitude of the vibration. The intensity depends on the energy per unity area of the wave.

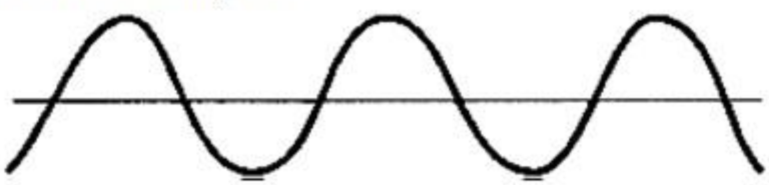
3.QUALITY OF SOUND

- The same note on different instrument sound differently. They differ in quality of sound or timber. The quality of sound depends on the number of frequency produced. The notes consist of main or fundamental frequency mixed with the overtones (the multiples of fundamental notes).

A note played on piano.



A note from tuning fork



FORCED VIBRATIONS AND RESONANCE

Forced vibrations are the vibration that occurs in a system as a result of impulses received from another system vibration nearby.

Example; when a tuning fork a sounded and placed on a bench or a hollow box, the sound produce is quite loud the box or bench is set into forced vibration by the vibration tuning fork.

Resonance is the phenomena where by the response of the system that a set into forced vibration when the driving frequency is equal to the natural frequency of the responding system.

NB: A resonance is said to occur when a body or system a set into vibration or oscillation at its own natural frequency as a result of impulses received from another system which is vibration at the same frequency.

Example

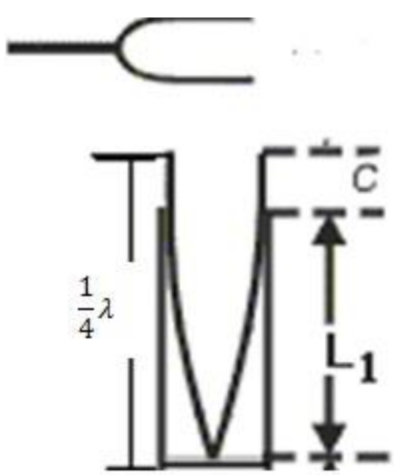
1. A group of troupes was marching towards bridge the bridge collapsed even before it s approached.

2. If a very loud sound is produced near the mouth of the glass bottle, the glass is likely to break.

3. The buildings are likely to collapse following the occurrences of the earth quake

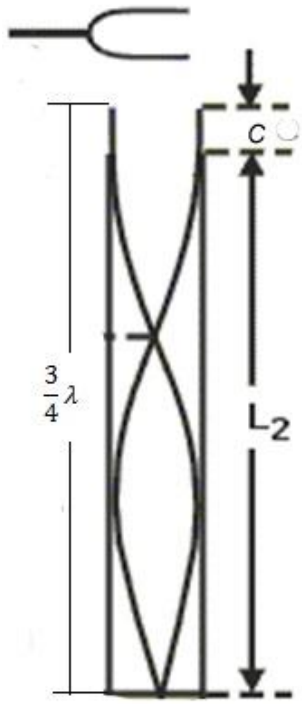
RESONANCE IN PIPES

When a turning fork is sounded at the top of a tube with one end open and the other closed, the air in the tube vibrate freely (resonates) at a certain length of a tube. The resonance is observed as a loud sound produced in the tube when the proper length obtained



$$L_1 + C = \frac{1}{4}\lambda \ldots\ldots\ldots\ldots (i)$$

FIRST OVERTONE



$L_2 + C = \frac{3}{4}\lambda \ldots\ldots\ldots\ldots (ii)$

Using equation (i) and (ii)

$L_2 + C = \frac{3}{4}\lambda$

$L_1 + C = \frac{1}{4}\lambda$

$(L_2 + C) - (L_1 + C) = \frac{3}{4}\lambda - \frac{1}{4}\lambda$

$L_2 - L_1 = \frac{1}{2}\lambda$

$2(L_2 - L_1) = \lambda$

By using wave equation

$V = \lambda f$

So; $V = 2f(L_2 - L_1)$

NB: The first resonance occurs when the air vibrates its fundamental frequency or first harmonic. The vibration at the open end of the pipe extends into the free air just above the open end of the pipe. The distance of extension is known as end- correction denoted by **C** or **e**. Thus the effective length of the pipe b is **L_1+C**

From **V = 2F(L_2 - L_1)**

V is the speed of sound in air column

f is frequency of sounds in air

Example:

1. The length of a closed pipe is 160mm. calculate the wavelength and the frequency of;

   i) The first overtone '

   ii) The third harmonic

Given that the speed of waves in air is 320m/s

$$0.16\text{m} = \frac{3}{4}\lambda$$

$$\frac{0.64}{3} = \frac{3\lambda}{3}$$

$\lambda = 0.213$

$V = \lambda f$

$f = \frac{v}{\lambda}$

$f = \frac{320}{0.213}$

f = 1502.34Hz

ii) $\frac{5}{4}\lambda = 0.16$

$\frac{5\lambda}{5} = \frac{0.64}{5}$

$\lambda = 0.124$

$V = \lambda f$

$f = \frac{v}{\lambda}$

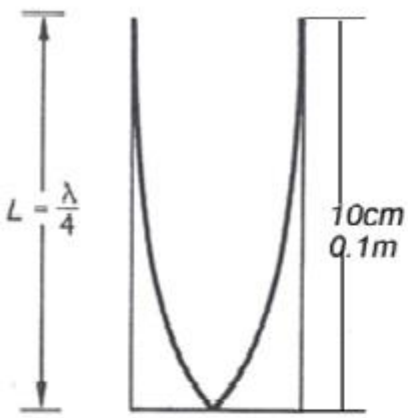
$f = \frac{320}{0.124}$

f= 2500Hz

2. A pipe closed at one and has a length of 100m. If the velocity of sound in air of the pipe is 340m/s. Calculate the frequency of;

a) The fundamental

b) The first overtone



$0.10 = \frac{1}{4}\lambda$

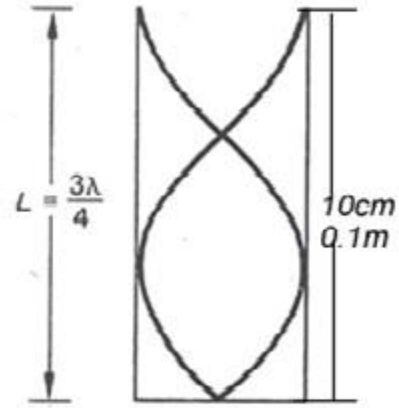
$0.4 = \lambda$

$f = \frac{v}{\lambda}$

$f = \frac{340}{0.4}$

$f = 850Hz$

$0.10 = \frac{3}{4}\lambda$



$0.4 = 3\lambda$

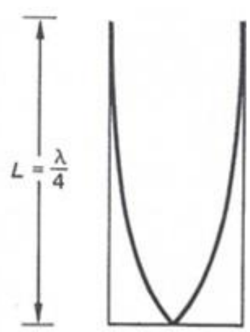
$\lambda = \frac{0.4}{3}$

$f = \frac{v}{\lambda}$

$f = \frac{340}{\frac{0.4}{3}}$

$f = 2550Hz$

First harmonic or fundamental note



$L = \frac{1}{4}\lambda$

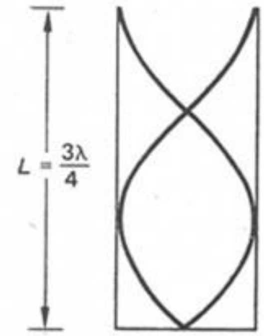
$\lambda = 4L$

$V = \lambda f_0$

$f_0 = \frac{v}{\lambda}$

$f_0 = \frac{v}{4L}$

2^{nd} harmonic or 1st overtone



$L = \frac{3}{4}\lambda$

$\lambda = \frac{4L}{3}$

$V = \lambda f_1$

$f_1 = \frac{V}{\lambda} = \frac{3v}{4L}$

$f_1 = \frac{3v}{4L}$

The comparison between f_1 and f_0

$\frac{f_1}{f_0} = \frac{\frac{3v}{4L}}{\frac{V}{4L}}$

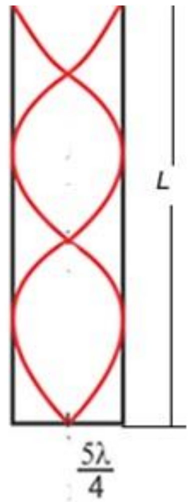
$\frac{f_1}{f_0} = \frac{3V}{4L} \text{ x } \frac{4L}{V}D$

$\frac{f_1}{f_0} = \frac{3}{1}$

$f_1 = 3f_0$

- Where 3 indicates the number of harmonics

2^{nd} overtone or 5^{th} harmonic



$$L = \frac{5}{4}\lambda$$

$$\lambda = \frac{4L}{5}$$

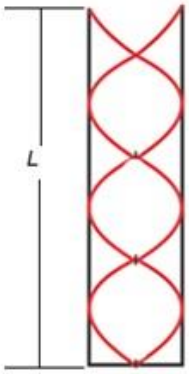
$$V = \lambda f_2$$

$$f_2 = \frac{v}{\frac{4}{5}\lambda}$$

$$f_2 = \frac{5v}{4L}$$

$$f_2 = 5f_0$$

3rd overtone or 7th harmonic



$$L = \frac{7}{4}\lambda$$

$$\lambda = \frac{4L}{7}$$

$$V = \lambda f_3$$

$$f_3 = \frac{v}{\lambda} = \frac{v}{\frac{4}{7}l}$$

$$f_3 = \frac{7v}{4l}$$

$$f_3 = 7f_o$$

Generally;

$f_0 = 1f_0$

$f_1 = 3f_0$

$f_2 = 5f_0$

$f_3 = 7f_0$

Where n = 0,1,2,3 …………… (Overtone)

Example 1; The speed of sound waves in air is found to be 340m/s. Find;

(a) The fundamental frequency

(b) The frequency of the 3rd harmonic

(c) The frequency of 9th harmonic

(d) The frequency of 51st harmonic

Given that the sound waves are probating in a closed pipe of length 700m.

Solution

(a) The fundamental frequency

$$f_0 = \frac{v}{4L}$$

$$f_0 = \frac{340}{4 \times 0.7}$$

$= 121.5Hz$

(b) The frequency of the 3rd harmonic

$$f_n = (2n + 1)f_0$$

$$f_3 = (2 \times 3 + 1)f_0$$

$$f_3 = 7 \times 121.5$$

$$f_3 = 850.5$$

(c) The frequency of 9th harmonic

$$f_n = (2n + 1)f_0$$

$$f_9 = (2 \times 9 + 1) \times 121.5$$

$$f_9 = 19 \times 121.5$$

$$f_9 = 2308.5$$

(d) The frequency of 51 harmonic

$$f_{51} = (2 \times 51 + 1)f_0$$

$$f_{51} = 103 \times 121.5$$

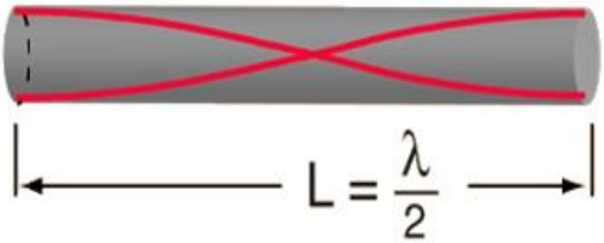
$$f_{51} = 12514.5$$

The frequencies are 12.5Hz, 850.5Hz, 2308.5Hz and 12514.5Hz

RESONANCE IN OPEN PIPES

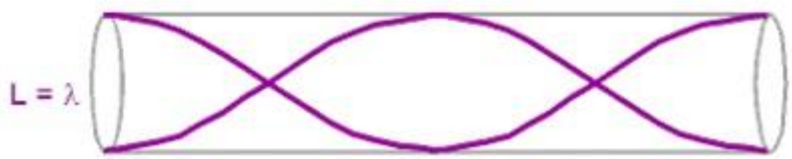
- Both ends are open

Fundamental note or first harmonic



$$\text{So} \quad L = \frac{1}{2}\lambda$$
$$\lambda = 2L$$
$$V = \lambda f_0$$
$$f_0 = \frac{v}{\lambda}$$
$$f_0 = \frac{v}{2L}$$

First overtone



So $L = \lambda$

$V = \lambda f_1$

$$f_1 = \frac{V}{\lambda} f_1$$
$$f_1 = \frac{V}{L}$$

Comparing f_1 and f_0

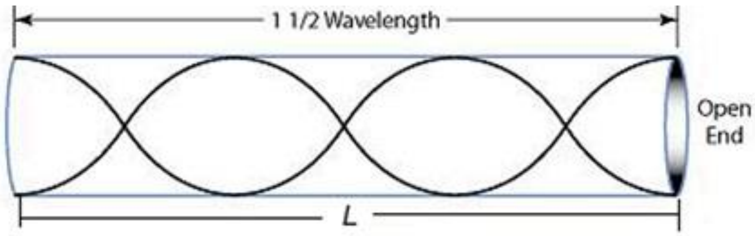
$$\frac{f_1}{f_0} = \frac{v}{L} \div \frac{v}{2L}$$

$$\frac{f_1}{f_0} = \frac{v}{L} \times \frac{2L}{v}$$

$$\frac{f_1}{f_0} = \frac{2}{1}$$

$$f_1 = 2f_0$$

Second overtone



So $L = \frac{3}{2}\lambda$

$\lambda = \frac{2L}{3}$

$V = \lambda f_2$

$f_2 = \frac{v}{\lambda}$

$f_2 = \frac{3v}{2L}$

$\frac{f_2}{f_0} = \frac{3v}{2L} \times \frac{2L}{V}$

$\frac{f_2}{f_0} = \frac{3}{1}$

$f_2 = 3f_0$ 3^{rd} harmonic

Third overtone

So $L = 2\lambda$

$$\lambda = \frac{L}{2}$$

$$f_3 = \frac{v}{\lambda}$$

$$f_3 = \frac{2v}{L}$$

$$\frac{f_3}{f_0} = \frac{2v}{2L} \times \frac{2L}{V}$$

$$\frac{f_3}{f_0} = \frac{4}{1}$$

$$f_3 = 4f_0 \quad 4^{rd}\text{harmonic}$$

Generally

$f_0 = 1f_0$

$f_1 = 2f_0$

$f_2 = 3f_0$

$f_3 = 4f_0$

$$f_n = (n + 1)\, f_0$$

Where $n = 0, 1, 2, 3\ldots$

Example;

1. Find the length of an open and air column required to produce fundamental frequency (first harmonic) of 480Hz. Take the speed of sound in air to be 340m/s.

From;

$$f_0 = \frac{v}{2L}$$

$$480 = \frac{340}{2L}$$

$$2L = \frac{340}{480}$$

$$L = \frac{340}{2 \times 480}$$

Length = 0.35m

2. Imani is playing an open end pipe. The frequency of the second harmonic is 880Hz. The speed of sound through the pipe is 530m/s.

•Find the frequency of the first harmonic and length of the pipe.

$f_1 = 2f_0$

$$f_1 = 2f_0$$

$$\frac{880}{2} = \frac{2f_0}{2}$$

$$f_0 = 440\text{Hz}$$

$$f_0 = \frac{v}{2L}$$

$$440 = \frac{350}{2L}$$

$$L = \frac{350}{440 \times 2}$$

L = 0. 398m

3. On a cold day Mathews blows a toy flute causing resonating in an open and air column. The speed of sound through the air column is 336m/s. The length of the air sound is 300m. Calculate the frequency of the $1^{st,}$ 2^{nd} , 3^{rd}, 4^{th}, 5^{th}, harmonics.

DATA GIVEN

V = 336m/s

L = 30cm=0.3m

From;

$f_0 = \frac{v}{2L}$

$f_0 = \frac{336}{2 \times 0.3}$

= 560Hz

The first harmonic is 560Hz

$f_1 = 2f_0$

=2 x 560

= 1120Hz

The second harmonic is 1120Hz

$f_2 = 3f_0$

= 3 x 560

= 1680Hz

The third harmonic is 1680Hz

$f_3 = 4f_0$

= 4 x 560

= 2240Hz

The forth harmonic is 2240Hz

$f_4 = 5f_0$

= 560

= 2800Hz

The fifth harmonic is 2800Hz

1. A flute is played with first harmonic of 196Hz. The length of the air column is 89.2cm. Find the speed of the wave resonating in the flute.

From;

$$f_0 = \frac{v}{2L}$$

$$196 = \frac{v}{2 \times 0.892}$$

V = 196 x 2 x 0.892

V = 349.664m/s

= 350m/s

The speed of the wave is 349.664m/s 350m/s

Revision Questions

1. A pipe closed at one end has a length of 10cm. If the velocity of sound in the air of the pipe is 340m/s. Calculate the frequency of;

   (a) The fundamental

   (b) 1^{st} overtone

Data given;

L = 10cm = 0.1m and V = 340m/s

a) $f_0 = \frac{v}{4L}$

$= \frac{340}{4 \times 0.1}$

= 850Hz

(b) fn = (2n +1) f_0

1. f_1 = (2x1+1) x 850

= 3 x 850

= 2550Hz

The fundamental note is 850Hz and the first overtone is 2550Hz

2. A pipe closed at one and has a length of 2.46m. Find the frequency of the fundamental and the first two overtones. Take 343m/s as the speed of sound in air.

i) L = 2.46m

V = 343m/s

$f_0 = \frac{v}{4L}$

$= \frac{343}{4 \times 2.46}$

= 34.85Hz

ii) fn = (2n+1) f_0

f_1 = (2 x 1 + 1) 34. 85

= 3 x 34.85

104. 55Hz

f_2 = (2 x 2 +1)34.85

= 5 x 34.85

174.25Hz

The frequency is 34.85Hz, 104.44Hz, and 174.25Hz

5. When a tuning fork of 512Hz is sounded at the top of the measuring cylinder which contains water. The first resonances are observed when the length of the air column (the distance from the mouth to the level of the water is 50Cm) and the second resonance is observed when the length of the air column (the distance from the mouth to the level of water) is 80Cm; using these observations. Calculate the velocity of water in air.

From;

$v = 2f(L_2 - L_1)$

$= 2 \times 512(0.8 - 05)$

$= 2 \times 512 \times 0.3$

307 .2m/s

The velocity of water in air is 307 .2m/s

WAVES Cont...

ELECTROMAGNETIC SPECTRUM

The electromagnetic spectrum is a continuous band of all electromagnetic waves arranged in order increasing or decreasing frequencies or wavelength change.

The electromagnetic spectrum includes wavelength such as;

1. Visible light
2. Infrared radiations
3. Radio waves
4. Ultraviolet radiations
5. X – rays
6. Gamma rays
7. Microwaves etc

NB: All the transverse waves traveling at the speed of **$3 \times 10^8 ms^{-1}$** and can travel through vacuum are the members of the electromagnetic spectrum. These waves are characterized by different frequency and wavelength.

The electromagnetic spectrum can be divided into seven major region or bands

Table 1.0 The electromagnetic spectrum

Wavelength(m)	Region(band)	Frequency(Hz)
$>10^{-1}$	Radio waves	$>3 \times 10^9$
$10^{-1} - 10^{-4}$	Microwaves	$3 \times 10^9 – 3 \times 10^{12}$
$10^{-4} \times 10^{-7}$	Infrared	$3 \times 10^{12} - 4.3 \times 10^{14}$
$7 \times 10^{-7} – 4 \times 10^{-7}$	Visible Light	$4.3 \times 10^{14} – 7.5 \times 10^{14}$
$4 \times 10^{-7} - 10^{-9}$	Ultraviolet Light	$7.5 \times 10^{14} – 3 \times 10^{17}$
$10^{-9} - 10^{-11}$	X-rays	$>3 \times 10^{17} – 3 \times 10^{19}$
$<10^{-11}$	Gamma rays	$>3 \times 10^{19}$

GAMMA RAYS (γ-rays)

- Gamma rays are the rays which have **short wave length** and **high frequency**. Gamma rays can be used to kill dangerous cell in humans but care is needed in their use, because they also attack and kill healthy cells. The gamma rays can be detected by **Geiger Miller Tube (G-Tube)** and **photographic plates** or **film**

X - RAYS

- X–Rays are produced when fast moving electrons are stopped by a metal target. X-Rays are used in **Radiology** (is the science of applying x-rays to medicine to produce pictures of internal organs in the body). They are also used to kill dangerous cells and tumors (abnormality) in the body it should be noted that healthy cell is also killed when the x- rays are used. They can be detected by **photographic plates**.

ULTRA VIOLET RADIATIONS (U V)

- Ultra violet radiations are just beyond the violet of the visible spectrum. They have longer wavelength than those of x-rays. UV radiations provide vitamin D from the sun. Excessive UV radiation can be harmful to the eyes and skin (recall albino). They are also used to detect **forged bank notes**.

VISIBLE SPECTRUM

- Visible spectrum is a narrow band of radiation of wavelength from pad to violet. The visible spectrum can be detected by human eye, light dependent resistor (LDR) such as photoelectric cell (solar cell and solar panels)

ROYGBIV

Where;

R=red

O=orange

Y=yellow

G=green

B=blue

I=indigo

V=violet

INFRARED SPECTRUM

- Infrared radiation is a band of radiations characterized by warmth. These heat radiations are produced from hot bodies such as, electric fires. The infra red radiations can be detected by **thermophile** and **thermometer**. All bodies emit infra red radiations.

MICROWAVES

1. The microwaves are produced by the oscillation of charges. These are used in cooking (oven) and in telecommunication, the beams of microwaves carry much more information that telephone wires.
2. Huge concave dishes are used to send and receive microwave signals.
3. The microwave are also used in the operation of RADAR (Radio waves detect and Ranging) and in TV waves

RADIO WAVES

1. Radio waves are produced by electrical oscillations in the circuit. They are used for communication purposes.

   NB: Lightning also produce radio waves. That is why radio programmers crackle when there is a thunderstorm. The radio waves have the **longest wavelength**.

Properties of electromagnetic spectrum

All forms of electromagnetic radiation have the following properties;

1. They are all transverse waves traveling in free space at 3 x10^8m/s.
2. They can be refracted , diffracted and exhibit the phenomena if interference.
3. They do not need material media for travel , they can even pass through vacuum.
4. They carry no electric charge.
5. They obey wave equation , $c=f\lambda$
6. They transfer energy from a source to a receiver in the form of oscillating electric and magnetic field.

BEAT FREQUENCY

- Beat frequency is a regular rise or fall of sound waves which have nearly equal frequencies.

  The beat frequency or the number of beats is given as the difference between two frequencies of sound.

  Beat frequency = $f_1 - f_2$ or $f_2 - f_1$

  Beat frequency = $|f_2 - f_1|$

Examples

1. Two sounds are emitted at the same time with frequencies 512Hz and 518Hz. Calculate the beat frequency.

   Solution

   from; $|f_2 - f_1|$

   Beat frequency $= |518 - 512|$

   $= |6|$

   $= 6\text{Hz}$

2. Given that the beat frequency is 4Hz. One of the sound waves has a frequency of 20Hz. Calculate the possible values of the frequencies of the other sound wave.

   Let the frequencies be f_1 and f_2

   Beat frequency $= |f_2 - f_1|$

   $4\text{Hz} = |f_2 - 20\text{Hz}|$

   $4\text{Hz} = f_2 - 20\text{Hz}$

   $4\text{Hz} + 20\text{Hz} = f_2$

24Hz = f_2

f_2 = 24Hz

4 = — |f_2 —20Hz|

4 = — f_2 +20Hz

4—20Hz = —f_2

—16Hz = —f_2

f_2 = 16Hz

The possible values are 24Hz and 16Hz

Question

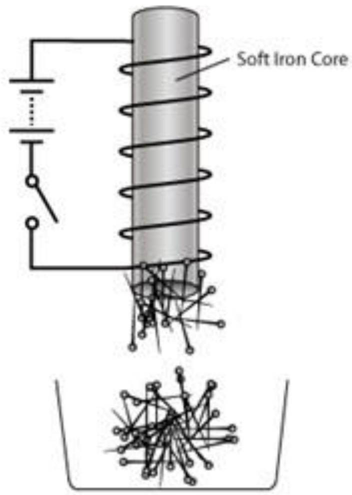
A 256Hz turning work produces sound at the same time with a, 249Hz. What is the beat frequency?

Ans: 7Hz

ELECTROMAGNETISM

INTRODUCTION

The phenomenon where by electricity creates magnetism is known as **electromagnetism**. When an insulated wire is wrapped with round an iron nail and the ends of the wire connected to the battery the nail becomes capable of picking up iron fillings paper. This is a simple electromagnet. The nail has been magnetized by the current in the wire. Disconnecting the wire from the battery the paper clips fall off. The nail looses most of first magnetism when the current is switched off. The passage of electric current along a wire creates a magnetic field around the wire. The lines of forces due to a straight current carrying wire are circles, center on the wire on the wire. The field is strongest near the wire the direction of the magnetic field is reversed if the direction of the current is reversed.

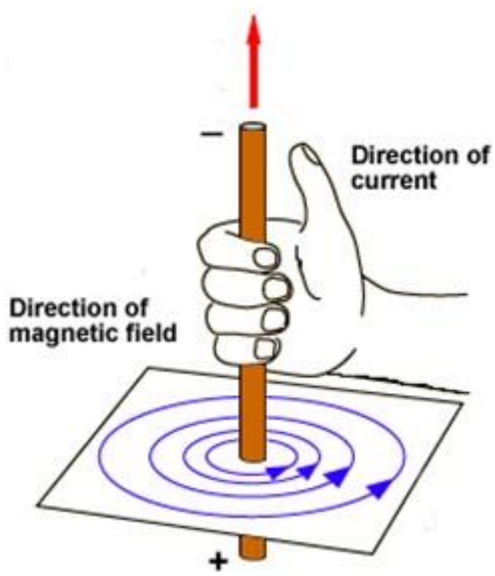


THE DIRECTION OF THE MAGNETIC FIELD

The direction of the magnetic field is determined by applying the following rules

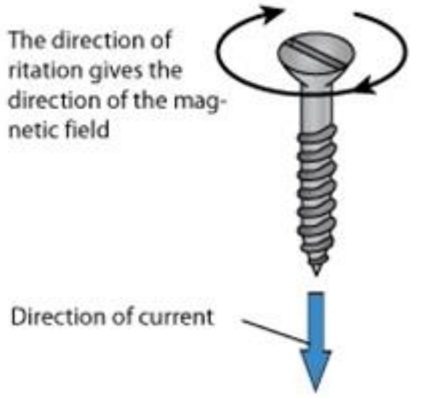
1. **Right hand grip rule**

States that if you grip a conductor in your right hand the thumb pointing in the direction of the current, the fingers point in the direction of the magnetic field



2. Maxwell's screw rule (The corkscrew rule)

States that "if a corkscrew is screwed in the direction of the current in the conductor, the screw rotates in the direction of the magnetic field"



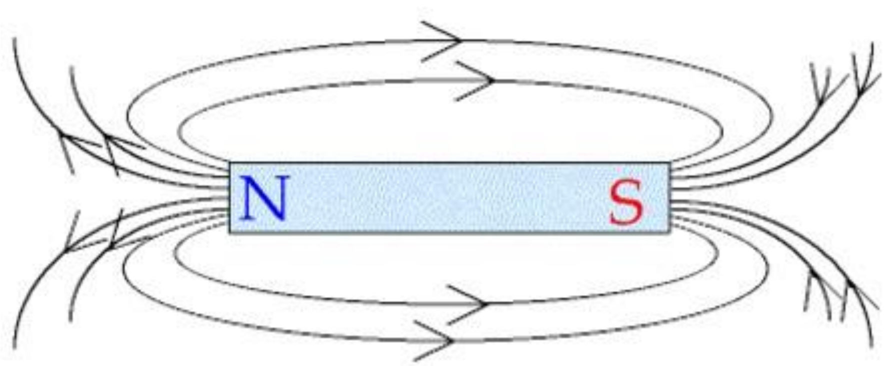
Maxwell's Screw Rule

THE DIRECTION OF THE MAGNETIC FIELD IN FIELD IN A SOLENOID

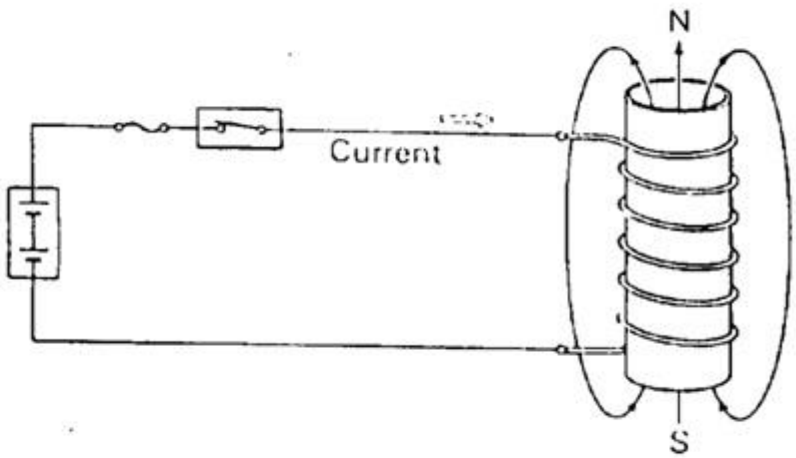
What is a solenoid?

A solenoid is a long coil of wire. The magnetic field lines created by current carrying solenoid are similar to that of a bar magnet.

Bar Magnet



However unlike a bar magnet the field line pass through the solenoid a long its axis.



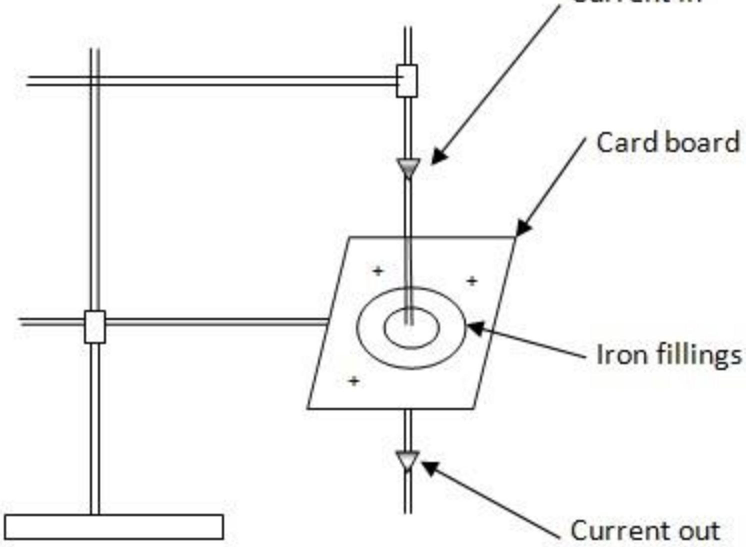
RULES

1. **Right hand grip rule**

   If you hold a wire in the right hand and your thumb lies parallel to the wire, the thumb then points in the direction of the flow of current . The remaining fingers point in the direction of the magnetic field.

2. **The corkscrew rule**

   If a right handed cork screw is turned so that its points travel along the current direction, the direction of rotation of the cork screw gives the direction of the magnetic field.



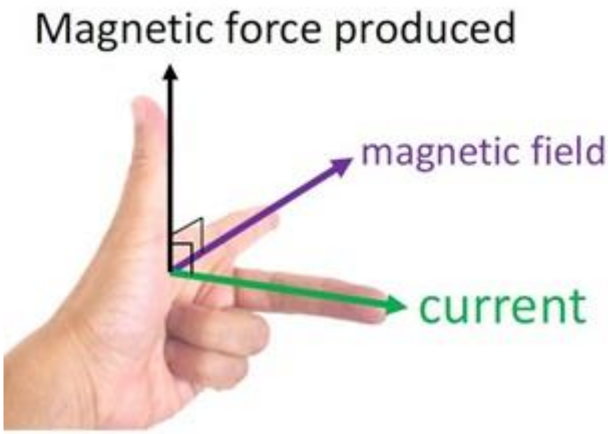
THE FACTORS AFFECTING THE STRENGTH OF A MAGNETIC FIELD AROUND A SOLENOID

1. The number of turns (windings)The more the number of turns the stronger is the magnetic field of the solenoid and vice verse

2. The amount of current flowing.
3. The bigger is the current flowing the stronger is the magnetic field and vice verse.
4. Placing a soft iron core along the axis of the solenoid increases the strength of the magnetic field.

FORCE OF A CURRENT CARRYING CONDUCTOR IN A MAGNETIC FIELD

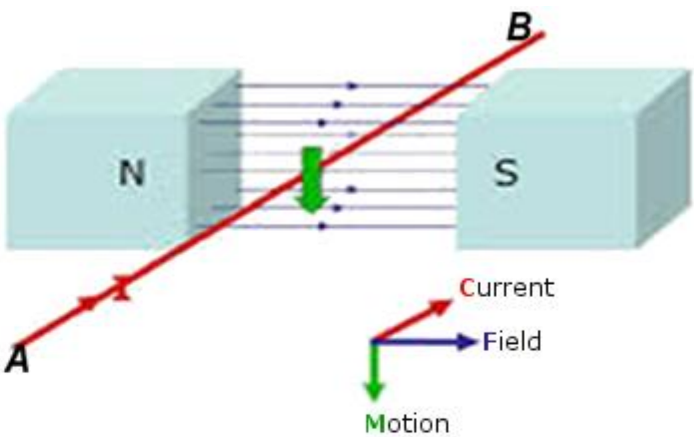
- The force on a conductor of length **(L)** carrying a current **(I)** placed perpendicular to the magnetic field is given by Fleming's left hand rule which states that , if you hold three fingers of your left hand perpendicular to each other such that the forefingers and second finger point in the direction of the magnetic field and the current respectively then the thumb point in the direction of the force (motion) acting on the conductor



Fleming's Left Hand Rule

Demonstration

Consider the arrangement below



NOTE;

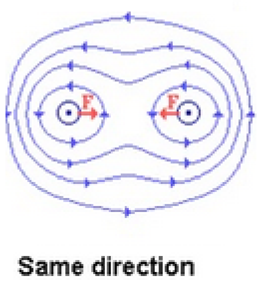
When the current I flows the conductor **AB** moves in the direction shown.

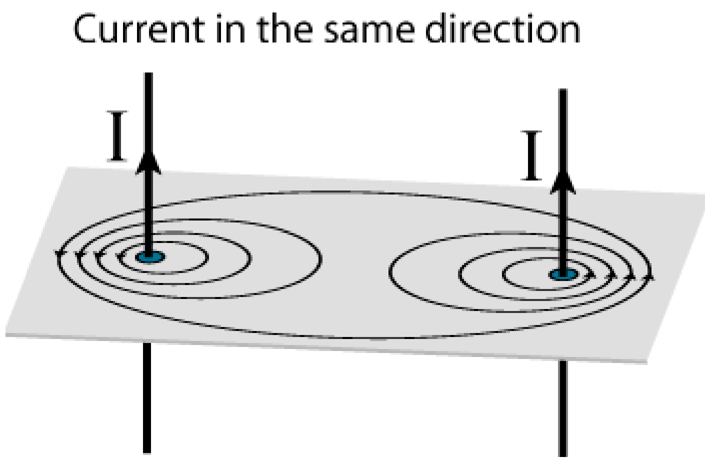
FORCE DUE TO TWO PARALLEL CURRENT CARRYING CONDUCTORS

- Consider two parallel current carrying conductors such that the some amount of current, **(I)** flows through the two conductors since each conductor is a carrying a current, it produces a magnetic field. Each conductor carries a current in the magnetic field of the in the other conductor. Hence each conductor experiences a force depending on the direction of the currents.

Case 1:

THE FORCE BETWEEN TWO CURRENT CARRYING CONDUTORS IN THE SAME DIRECTION



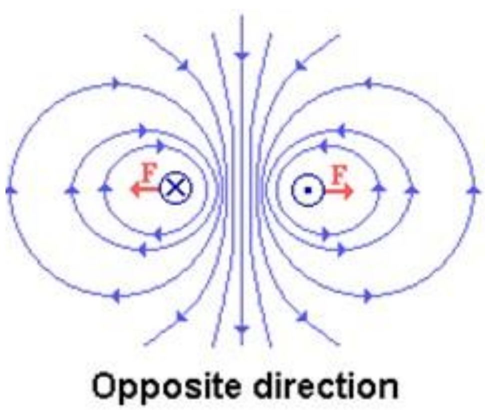


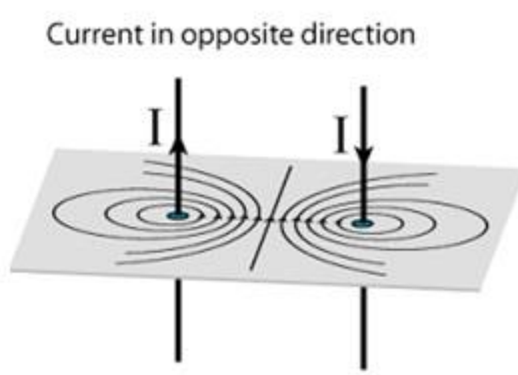
Observation

- The strips of the conductors attract each other when the currents are flowing in the same direction (because the forces acting on the wires).

Case 2

THE FORCE BETWEEN TWO CURRENT CARRYING CONDUCTORS IN OPPOSITE DIRECTION





Observation

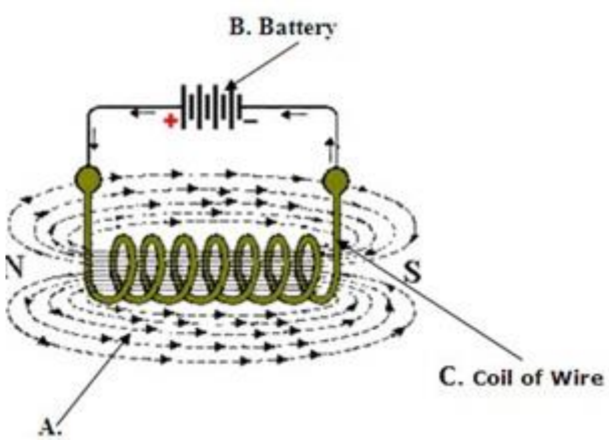
The strips of the conductor repel each other where the currents are flowing in the opposite direction.

THE ELECTROMAGNET

• An electromagnet is a coil of wire wound on a soft iron core controlled by an electric current. When the electric current flows in the iron core the soft iron core is magnetized. When the current is switched off, the magnetism of the electromagnet disappears as it was magnetized temporally, unlike that of a permanent magnet. The soft iron core is magnetized only when the current flows in the coil.

NB;

The poles of the electromagnet are by Maxwell right hand grip rule.



NOTE;

By clock Rule:

When viewing the end of the solenoid, if the current flows in a clockwise direction the end of the soft iron bar is South Pole. **(S)** If the current flows in a (**N**) if clockwise direction the end of the soft iron is North Pole **(N).**

THE DAILY APPLICATIONS OF ELECTROMAGNETS

- The electromagnet is an essential part of many electrical devices including the following;

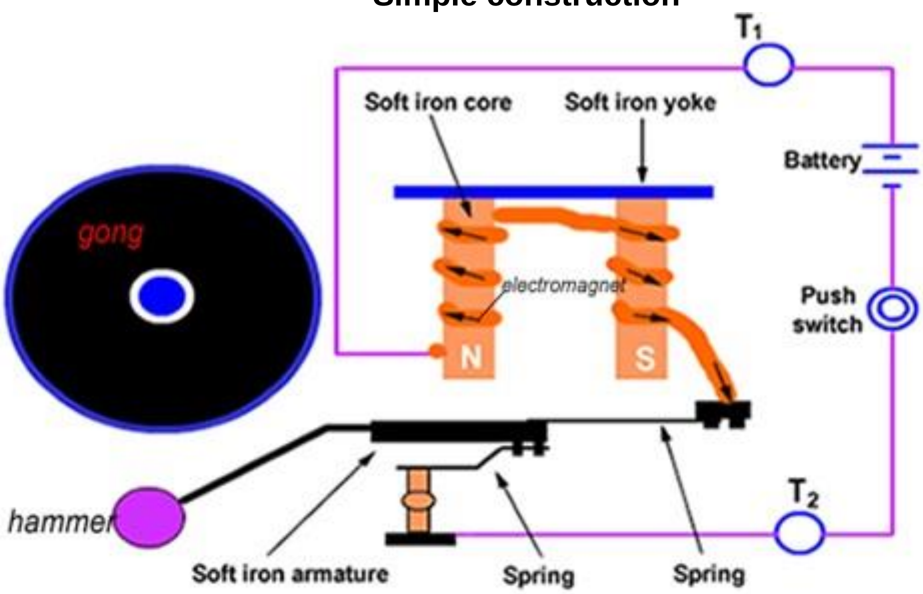
1. Electric bell

2. Telephone receiver
3. Relay doors
4. Scrap yards
5. Magnetic tap

1. **AN ELECTRIC BELL**

- An electric bell is a bell which uses electricity to function or operate.

Simple construction



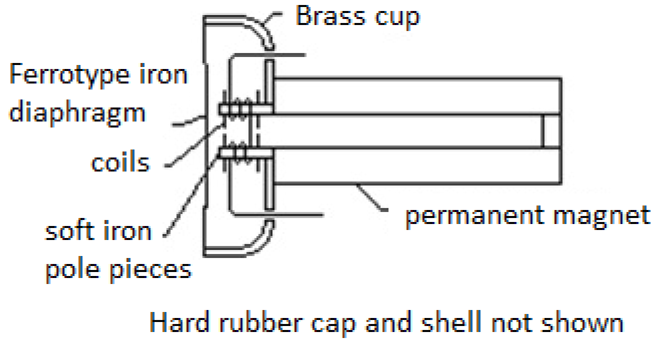
Mode of action

When the bell is connected to the power supply the soft iron armature is pulled onto the electromagnet the hammer will strike the gong (hits the gong) and the sound will be heard. This action opens the screw contact and the electromagnet switches off. When the electromagnet is switched off; the soft iron armature returns back to close the screw contact. Again and so the whole cycle repeats itself continuously.

1. **THE TELEPHONE RECEIVER AND LOUD SPEAKER**

- The telephone receiver contains a U-shaped magnet formed by placing a short permanent magnet across the ends of the two soft iron bars to exert a pull on a spring magnet alloy diaphragm.

Simple construction



Telephone Receiver

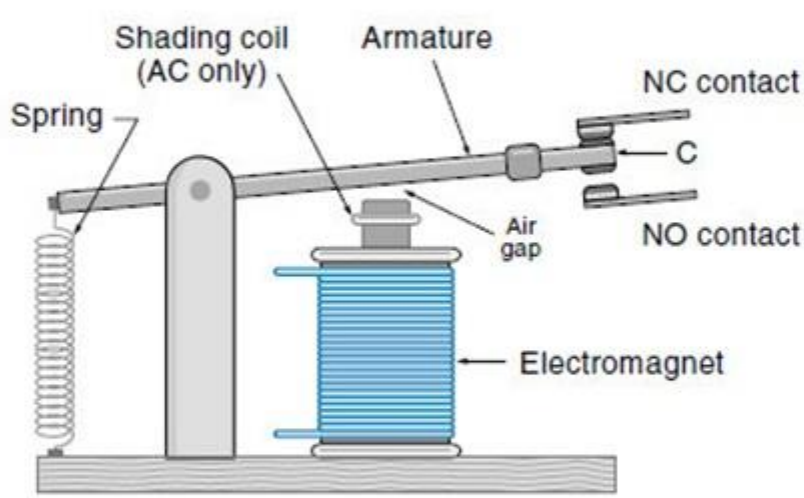
Mode of action

When a varying current is posed through the coil, the magnetic field of the permanent magnet produce a corresponding vibration in the pull of the diaphragm therefore vibrates and reproduces a copy of the sound waves which entered the microphone.

- The electromagnet is used in scraping to lift car bodies or heavy loads. The iron core of the electromagnet must lose its magnetism to release its load when the current is switched off.
- The electromagnets are also used in relays systems and automatic switches.

Simple construction of automatic switches

An electromagnetic relay.



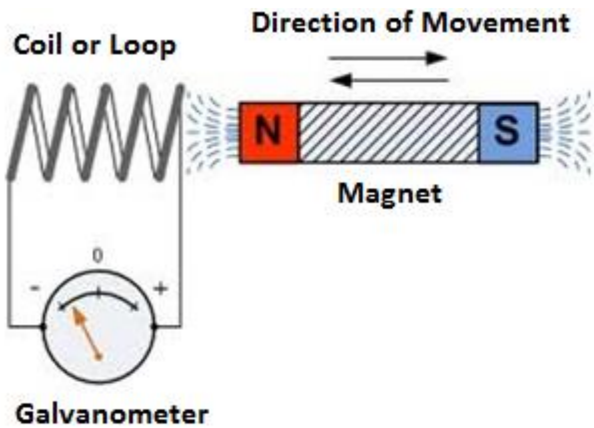
(a) Parts of the relay

NB: The relay system or automatic switches are used in control circuits to switch the machines on and off. When the current passes through the electromagnet, the iron armature is pulled onto the electromagnet the armature turns about the pivot and closes the switch gap. The relay system or automatic switches are used in making the relay doors or automatic doors in modern offices or banks.

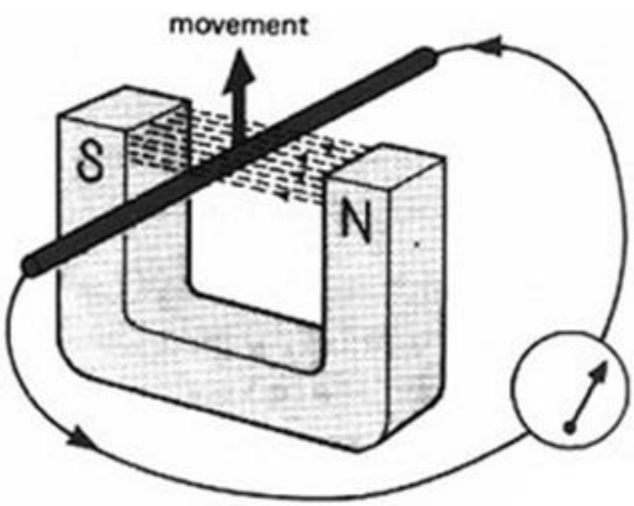
ELECTROMAGNETIC INDUCTION

- Whenever there is relative motion between the magnetic field and the conductor, the electromotive force is induced. The lines of magnetic field are sometimes referred to as magnetic flux. An electromotive force is induced whenever the magnetic flux linking a conductor is cut.

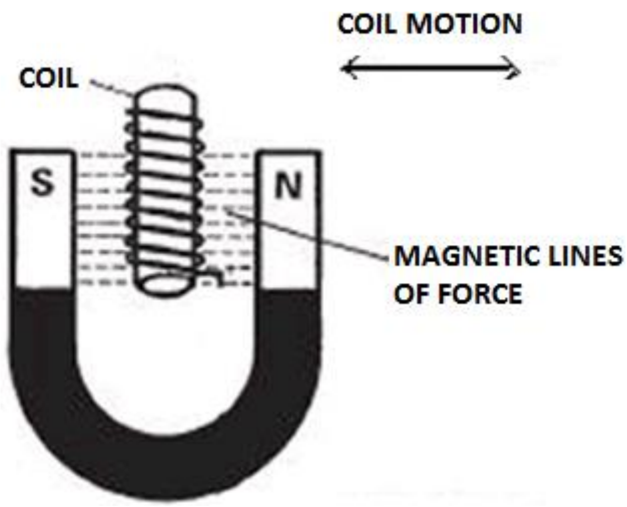
DEMONSTRATION



If a magnet is inserted into a solenoid connected to a galvanometer the current is observed to flow. This process is known as **electromagnetic induction**. The deflection observed on the pointer of the galvanometer proves that there in is an induced current flowing in the conductor when cut by the magnetic flux.



Deflection is observed



no deflection is observed

THE LAWS OF ELECTROMAGNETIC INDUCTION

FARADAY'S LAW

States that, "the size of the induced electromotive force a directly proportional to the rate at which the magnetic flux are cut by the conductor".

THE FACTORS AFFECTING THE MAGNITUDE OF THE INDUCED EMF

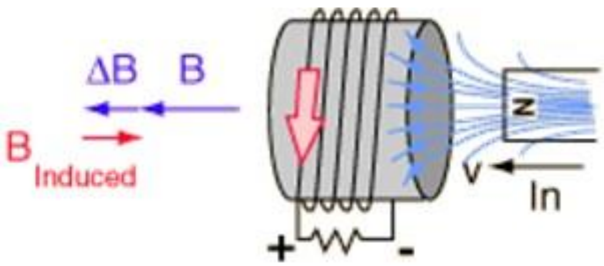
1. The magnitude (strength) of the magnetic field.   The stronger is the magnetic filled, the bigger is the inducing electromotive force and vice verse.
2. The speed of the relative motion between the magnetic field and the conductor. The higher the speed at which the magnetic fluxes are cut by a conductor, the bigger is the inducing electromotive force and vice verse.
3. Number of turns of the coil. The more the coils the higher is the inducing emf and vice verse.

LENZ’S LAW

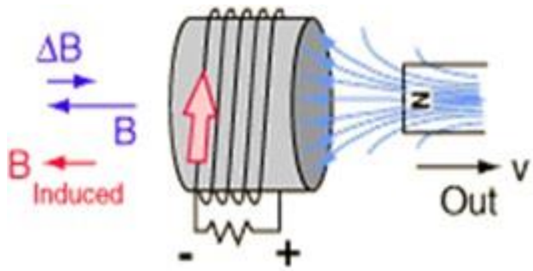
States that; "the direction of the induced electromotive force is such as to oppose the effect which is producing it".

NB: Lenz’s law can be explained by considering a magnet moving into a coil.

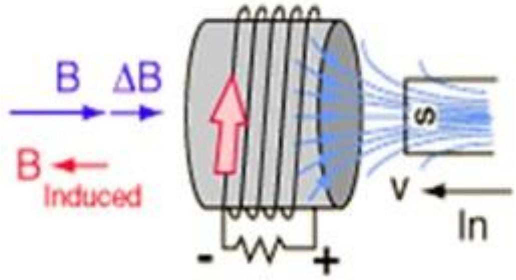
1. When the north pole of a magnet is moved into the coil.



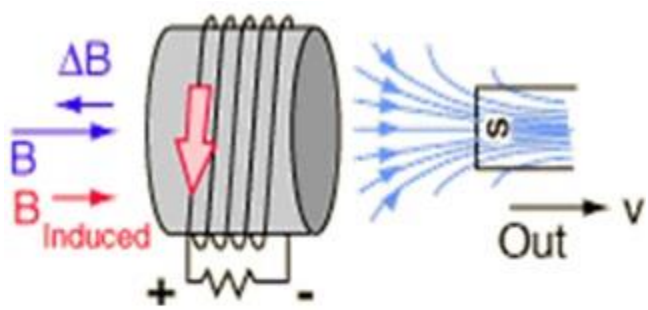
When the north pole is moved away from the coil.



When the south pole of a magnet is moved into the coil.



When the south pole is moved away from the coil.



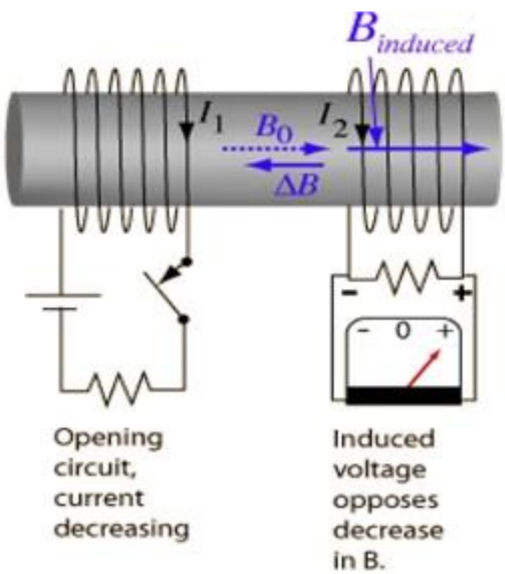
NOTE:

- When a north pole of a magnet is moved into a coil the magnetic flux increases, this induces electromotive force in the coil. The induced electromotive force results into induced current which flows in the direction to oppose this increase in magnetic flux.

SELF INDUCTION AND MUTUAL INDUCTION

1. **Self induction**

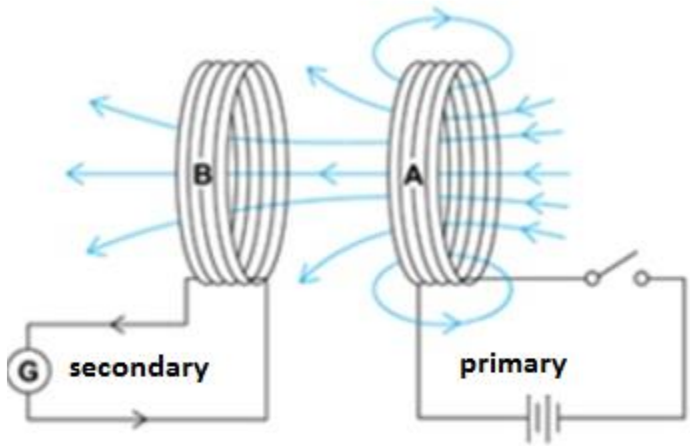
Suppose that the current flowing through a conductor varies with time, it creates a varying magnetic field that cuts across the conductor itself. The induced electromotive force in the conductor flows in the opposite direction to the original voltage. This is called **back electromotive force**.



Self induction therefore is the production of electromotive force in a conductor when the current flowing through it varies with time (Increases and decreases)

1. **Mutual induction**

Suppose that two coils are place near to each other a varying current in the first coil induces the current in the second coil.   This process is called mutual induction.



NB;

If the current in coil 1 (primary) is increasing, this produces an increasing magnetic flux in the coil 2 (secondary coil). Thus an induced electromotive force in the secondary coil produces a current that in turn produces magnetic flux in opposition to primary coil.

If the current in the primary coil decreases, this produces a decreasing magnetic flux in the secondary coil thus an induced electromotive force in the secondary coil produces a current that in turn produces a magnetic flux in the same directions as the primary coil.

NOTE: Mutual induction therefore is the production of electromotive force in one conductor when the current flowing in another conductor varies with time (increase and decreases).

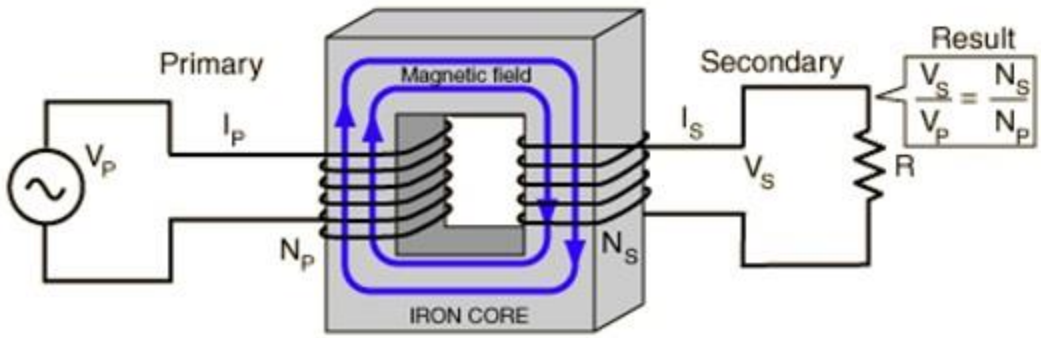
APPLICATIONS OF MUTUAL INDUCING IN PRACTICE

The phenomenon of mutual induction is applied in many areas. This include;

- Transformer
- Induction coil

• THE TRANSFORMER

The most common and useful application of the production of an induced electromotive force by varying the current in another coil is the transformer. The transformer depends on the use of A.C (alternating current). The center of the transformer consists of a terminated core of magnetic material which is easily magnetized and demagnetized. Around this core a primary coil and secondary coil are wound.

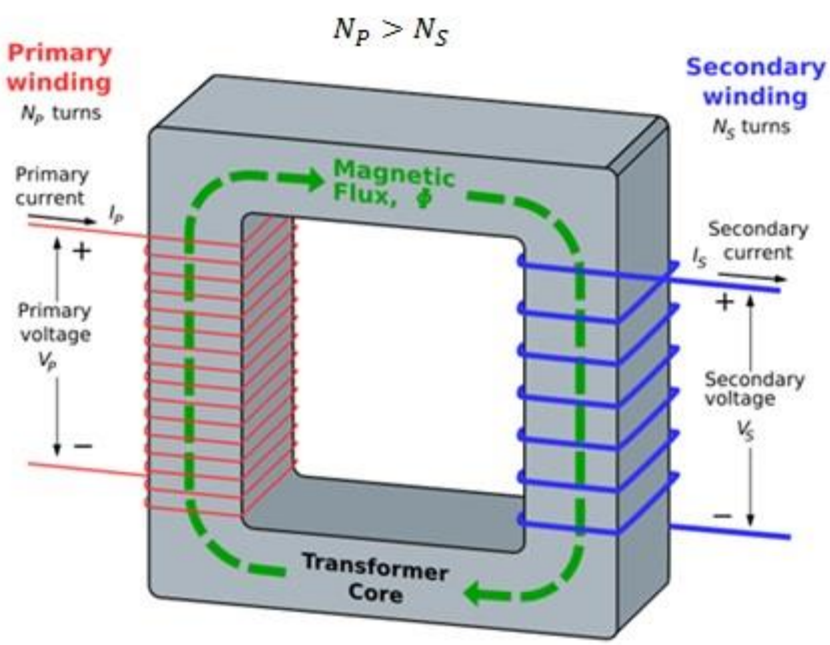


TYPES OF TRANSFORMER

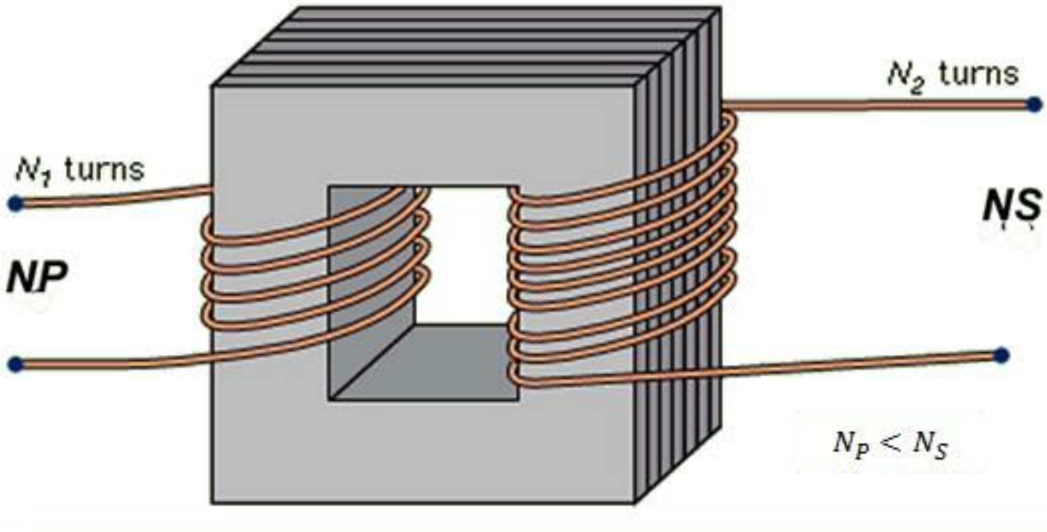
There are two types of transformers:

1. Step-down transformer
2. Step up transformer

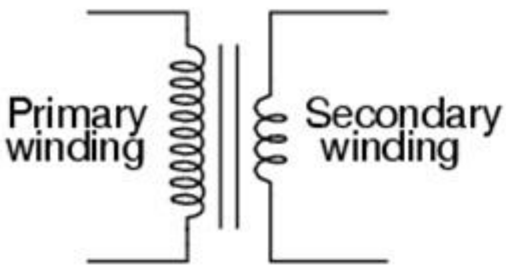
1. Step down transformer



2. Step up transformer

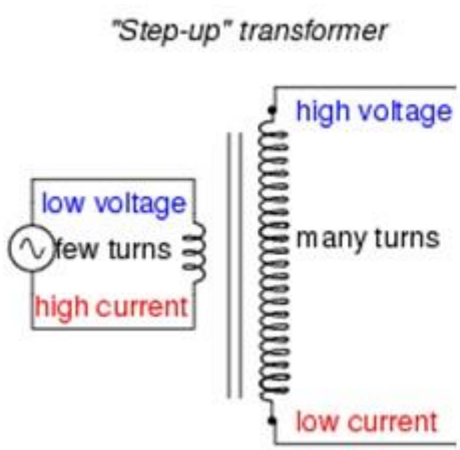


Alternatively;



$N_P > N_S$

STEP DOWN TRANSFORMER



$N_P < N_S$

STEP UP TRANSFORMER

Mode of Action

When an alternating electromotive force is applied to the primary coil an alternating magnetic field is produced in the core. This in turn, passes through the secondary coil and induces an alternating electromotive force in each coil. Since the magnetic in each will exist the same it follow that the electromotive force in each coil is directly proportional to the number of turns in the coil. Hence by using Faraday law of electromagnetic of induction.

$$Vp \propto Np \text{ and } Vs \propto Ns$$

$$\text{Thus } Vp = K\,Np$$

$$Vs = K\,Ns$$

$$K = \frac{Vp}{Np}$$

$$K = \frac{Vs}{Ns}$$

Thus;

$$\boxed{\frac{Vp}{Np} = \frac{Vs}{Vs}}$$

Also

$$\frac{Vp}{Vs} = \frac{Np}{Ns}$$

The above equation is known as Transformer Equation

NB; If $V_P > V_S$, the transformer is step down and if, the transformer is step up.

The efficiency of a Transformer

• As much energy can be obtained from the secondary coil as is put into the primary coil, the transformer is said to operate with high efficiency.

In practice a transformer cannot operate with an efficiency of 100% part of the energy is wasted as heat sound and part of the energy is wasted in overcoming the friction between the moving parts of the transform.

Hence;

$$\text{efficiency} = \frac{\text{power output}}{\text{power input}} \times 100\%$$

But;

$\text{power out put} = \text{Is} \times \text{Vs}$

$\text{power inpute} = \text{Ip} \times \text{Vp}$

Thus;

$$\text{efficiency} = \frac{\text{IsVs}}{\text{IpVp}} \times 100\%$$

For practical transformer,

$\text{IsVs} < \text{IpVp}$

hence e < 100%

For a perfect transformer;

Power output = power input

$\text{IsVs} = \text{IpVp}$

$\frac{\text{Is}}{\text{Is}} = \frac{\text{Vp}}{\text{Vs}}$

$\text{Vp} = \text{IpR}$

$\text{Vs} = \text{IsR}$

Also;

Vp=IpR
Vs=IsR

NOTE: From transformer equation

$\frac{\text{Vp}}{\text{Np}} = \frac{\text{Vs}}{\text{Ns}}$

But , Vp=IpR, Vs=IsR

$$\frac{IpR}{Np} = \frac{IsR}{Ns}$$

$$\frac{Ip}{Np} = \frac{Vs}{Ns}$$

Also; $\frac{Ip}{Is} = \frac{Np}{Ns}$

Also; $\frac{Is}{Ip} = \frac{Ns}{Np}$

Examples:

1. A transformer is used to step down 120V mains to 24volts, for kitchen use. If the primary coil has 400 turns , find the number of turns is the secondary coil

   Data:

$$\frac{Vp}{Np} = \frac{Vs}{Ns}$$

$$\frac{120v}{400} = \frac{24}{Ns}$$

$$Ns = \frac{400 \times 24}{120} \text{ turns}$$

V_p = 120V

V_s = 24V

N_p = 400 turns

N_s = ?

Using transformer equation

Ns = 80 turns

- The number of turns is the secondary coil is 80

2. A step up transformer has 5000 turns in the secondary coil. And 500turns through

the primary coil. An alternative current of 5A flows in the primary coil when connected to a 12V A.C supply.

(a) Calculate the voltage across the secondary coil.

(b) If the transformer has an efficiency of 90% what is the current in the secondary coil?

Data given;

N_s = 5000

N_p = 500

V_p = 12

I_p = 5

V_s = ?

From transformer equation

$$\frac{Vp}{Np} = \frac{Vs}{Ns}$$

$$\frac{12}{500} = \frac{Vs}{5000}$$

$$Vs = \frac{5000 \times 12}{500} \text{volt}$$

- The voltage across the secondary coil is 120V

efficiency=90%
Ip=5
Vp=12
Vs=120
Is=?

From;

$$\text{efficiency} = \frac{IsVs}{IpVp} \times 100\%$$

$$90\% = \frac{Is \times 120}{5 \times 12} \times 100\%$$

$90/100=2Is$

$0.9=2Is$

$Is=0.45A$

- The current in the secondary coil is 0.45A

3. A step down transformer is used to light a 12V, 24W lamp from 240 volts mains. The current through the primary coil is 125mA.

What is the efficiency of the transformer?

$$\text{efficiency} = \frac{IsVs}{IpVp} \times 100\%$$

$$\text{efficiency} = \frac{2\times12}{0.125\times240} \times 100\%$$

$= 80\%$

- The efficiency of the transformer is 80%

CLASS QUIZ

1. A transformer is used to step down 24V mains supplier to 12V for laboratory use, if the primary coil has 600turns. Find the number of turns in the secondary coil.

   Data

   $V_p = 240V$

   $N_p = 600$

   $Vs = 12V$

   $Ns = ?$

   Using transformer equation

$$\frac{Vp}{Np} = \frac{Vs}{Ns}$$

$$Ns = \frac{12\times600}{240}\text{turns}$$

$Ns = 30$

- The secondary coil has 30 turns

2. A current of 0.6A is passed through a step up transformer with a primary coil of 200 turns. A current of 0.1 A is obtained in the secondary coil. Find the number of turns in the secondary coil and the voltage across if the primary coil is connected to 240V mains

1. **Data**

$I_p = 0.6$

Is = 0.1

$N_p = 200$

Ns = ?

Using transformer equation

$$\frac{Vp}{Np} = \frac{Vs}{Ns}$$

$$Ns = \frac{0.1 \times 200}{0.3} \text{turns}$$

Ns = 33.3

- The secondary coil has approximately 33 turns

$$\frac{Vp}{Np} = \frac{Vs}{Ns}$$

$$\frac{240}{200} = \frac{Vs}{33}$$

$$Vs = \frac{240 \times 33}{200}$$

Therefore the voltage a cross the secondary coil is 39.6V

3. A step up transformer has 10000 turns in the secondary coil and 100 turns through the primary coil. An alternate current of 5.0A flows through the primary circuit when connected to 12V alternate current Supply

(a)Calculate the voltage a cross the secondary coil

(b)If the transformer has an efficiency of 90% what is the current in the secondary coil?

Data

Ns = 10.000

N_p = 100

I_p = 5.0A

V_p = 12
Vs = ?
Using transformer equation

a) $\frac{Vp}{Np} = \frac{Vs}{Ns}$

$$\frac{12}{100} = \frac{Vs}{10000}$$

$$Vs = \frac{12 \times 10000}{100}$$

Vs = 1200Volts

- The voltage a cross the secondary coil is 1200V

$$\text{efficiency} = \frac{IsVs}{IpVp} \times 100\%$$

$$\frac{90}{100} = \frac{Is \times 1200}{5 \times 12}$$

$$\frac{0.5}{20} = \frac{Is \times 20}{20}$$

Is = 0.045

- The current in the secondary coil is 0.045 A

3. A low voltage outdoor lighting system uses a transformer to step down a 240 voltage house hold voltage to 24 voltages. The lighting system has 6 lamps with a total resistance of 10Ω

(a) What is the current in the secondary coil of the transformer

(b) What is the current in the primary coil

1. **Data;**

   Vs = 240V

   Vs = 12V

   R = 10Ω

   From V = IR

   240 = I x 10

   $$I = \frac{240}{10}$$

   I = 24 A

- The current in the primary coil is 24 A

From V = IR

240 = I x 10

$$I = \frac{24}{10}$$

I = 2.4 A

The current in the secondary coil is 2.4 A

5. The ratio of the number of in the secondary coil in a transformer to that in the primary coil is 16:1. If the current in the secondary circuit is 4.0A. What is the current in the primary circuit?

From

$$\frac{Ip}{Np} = \frac{Vs}{Ns}$$

$$\frac{Ns}{Np} = \frac{Is}{Ip}$$

$$\frac{16}{1} = \frac{4}{Ip}$$

$$Ip = \frac{4 \times 1}{16}$$

Therefore the current in the primary coil is 0.25A

DAILY USES OF TRANSFORMERS

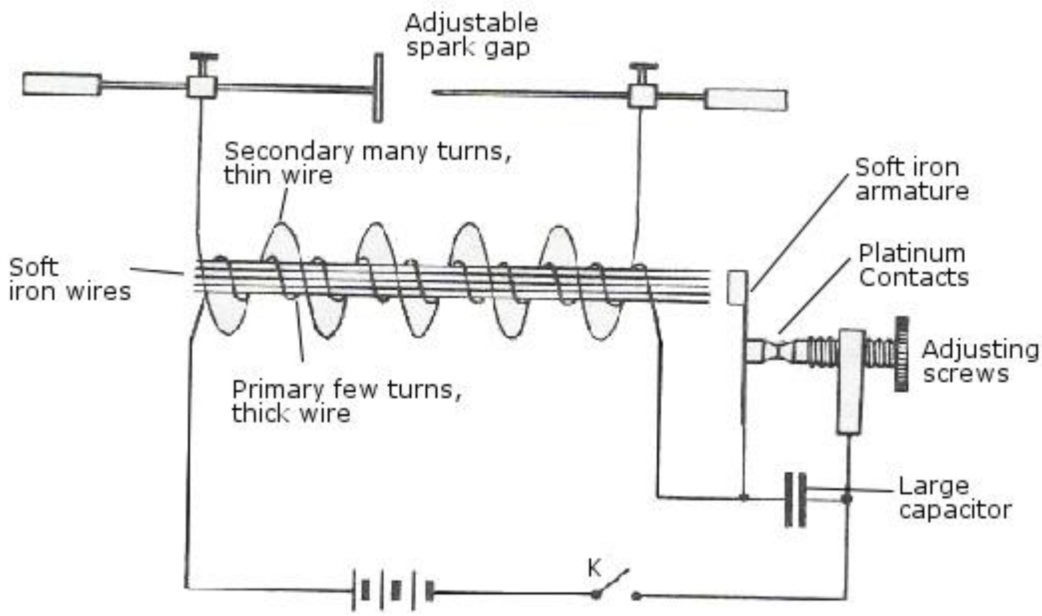
1. The transformer is used in power stations to step up voltage for transmission from the station into the areas of consumptions.

2**.** The transformers are used to step down the voltage to the area of consumption as for as the value of voltage required for domestic use is connected.
3. Electricity is transmitted from one part of the country to another part by grid system.

AN INDUCTION COIL

- An induction coil is an electrical instrument which consists of two coils (primary coil and secondary coil) wound one over the other on an iron core. An induction coil is used to produce high voltage **alternating current (ac)** from low voltage **direct current (dc)**. The primary coil has fewer numbers of turns than the secondary coil which is wound on top of the primary coil.

Structure of induction coil



Mode of action

The current in the primary circuit is switched on and off by a make break mechanism. This produces changes in current and magnetic field which are necessary for electromagnetic induction to occur in secondary coil. When the current in the primary coil is switched on the induced magnetism in the iron core attracts the soft iron armature which creates a gap between the two contacts (which break the primary coil circuit). This switched off the current. As the induced magnetism disappears away, the soft iron armature springs back, closes the contacts and completes the circuit gain. This allows the current to flow in the primary coil again. This process repeats itself automatically. Every time the current is broken the magnetic flux in the core which links the secondary coil decreases suddenly so that an electromotive force is induced in the secondary coil. Since the secondary coil, has very large number of turns as compared to the primary coil, the induced electromotive force is sufficiently high to causes the spark across the adjustable spark gap. A large capacitor **C** prevents small charge which is likely to form a cross the contacts **B** due to self induction of the primary coil. When the soft iron armature vibrates to and from, a process switches the current on and off hence alternating current **(ac)**

THE USES OF THE INDUCTION COIL

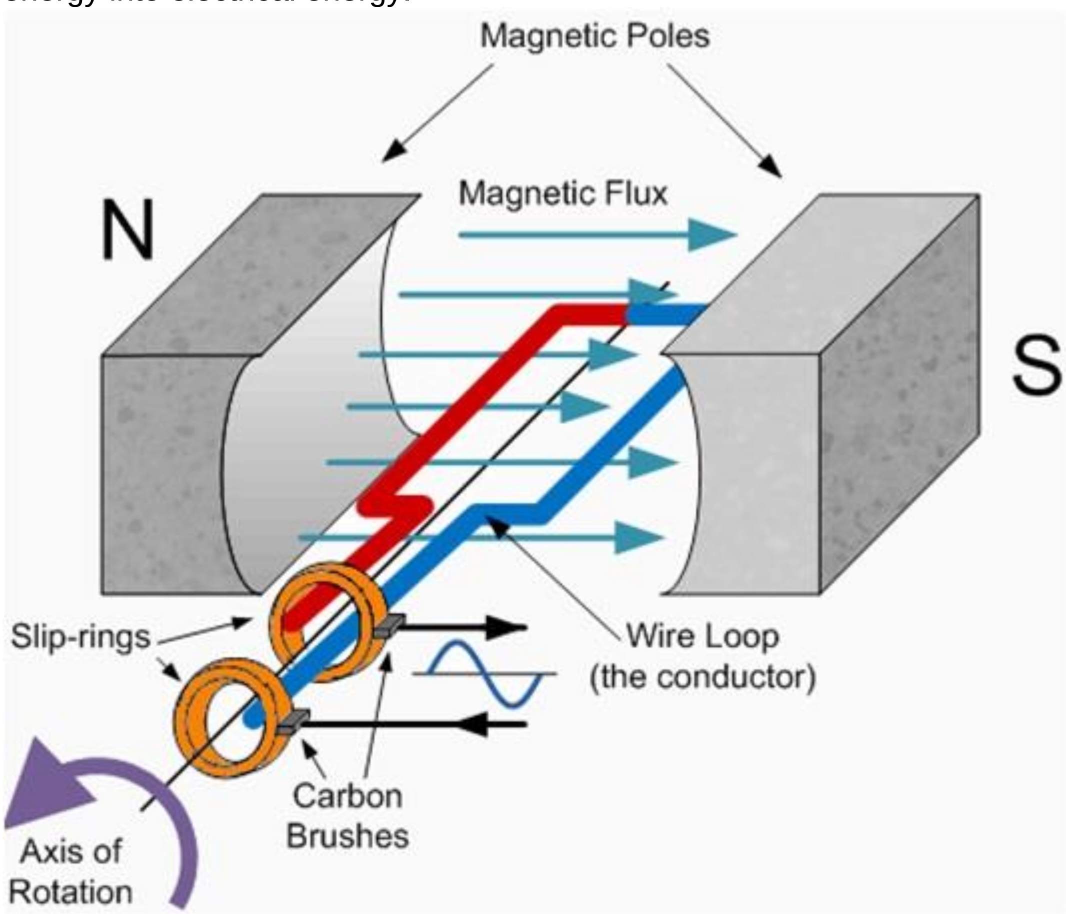
1. The induction coil is used in the ignition system of the internal combustion engines e.g. Petrol engine and diesel engine.

2. An induction coil is used to trigger the tubes in cameras (depending on the lightning conduction).
3. An induction coil is used for operating X –ray tube.

ALTERNATE CURRENT GENERATOR

A DYNAMO

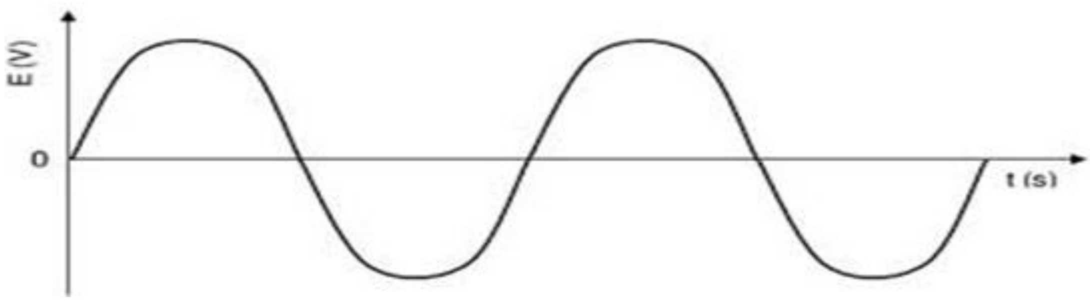
- A generator or a dynamo is an electrical device which converts mechanical energy into electrical energy.



Simple structure

Action

- By Fleming's left hand rule when the coil **CDEF** is rotated between the poles of a magnet an inducing electromotive force is created. The induced current flows to an external circuit when the coil is connected by slip rings **A_1** and **A_2** which are mode of copper**.** **A_1** is connected to side **CD** of the coil and **A_2** is connected to side **EF** of the coil. **A_1** is in contact with brush **B1** and **A_2** is in contact with Brush, **B_2**. When the side **CD** moves upwards the current flow from **C** to **D** and hence **E** to **F.** Half a revolution later**, EF** will be in the position previously occupied by **CD** the direction of the current is reversed i.e. From **F** to **E** and **D** to **C**. The direction of the induced electromotive force and the current change every half revolution. The magnitude of the induced electromotive force varies as the coil rotates , hence alternatively current.



- The maximum value of the inducing electromotive force is known as **peak value.**
- The time taken for one revolution , **T** is called **period** while the frequency of the generator is calculated from

$f = \frac{1}{T}$ (the number of complete revolutions mode per second).

NB: The induced electromotive force is a maximum value when the coil is perpendicular to the magnetic flux and the induced electromotive force is a minimum value when the coil is horizontal.

DAILY APPLICATIONS OF A.C. GENERATOR

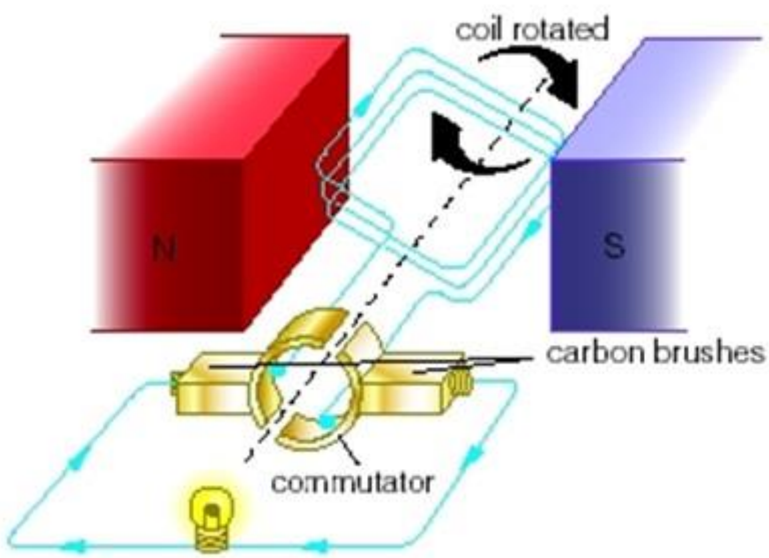
1. Most generators are designed to produce alternate current. The generators are often called alternators. In the car alternate the armature is rotated by a belt which is moved by a crank of the engine.

- In H.E.P, large alternators generate the alternate current mains electricity, the energy to rotate such armature is obtained from sources like falling water from a dam , compressed from steam, from a nuclear reactor or from steam underground (Geothermal electricity).

D.C. GENERATOR

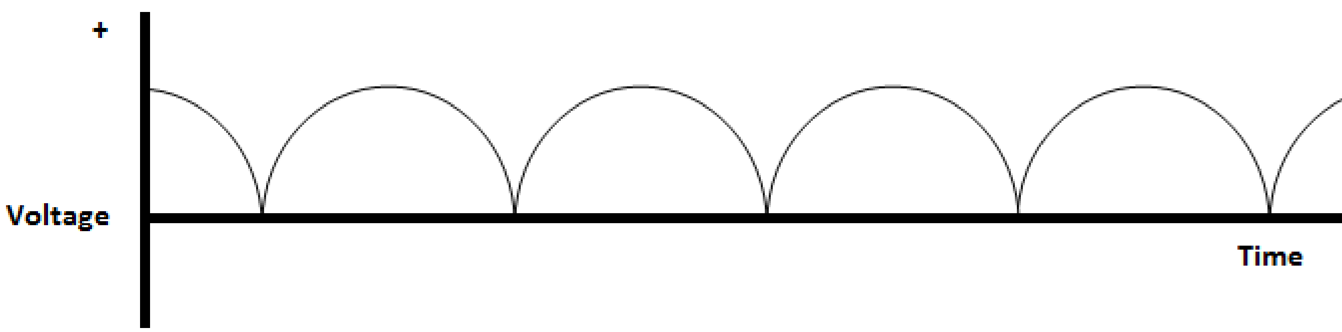
- A direct current generator is constructed by replacing the slip rings in alternate current generator with commutator.

Simple construction



Action

The two carbon brushes **B_1** and **B_2** are connected to the outsides circuit resting against the rotating communication. The commutators rotate with a loop of wire. When the loop is rotated in the magnetic field, the induced electromotive force creates. However after a rotation of **180^0,** the connections to the external circuit reverse but the direction of the current does not alter. Hence direct current.



MOVING COIL GALVANOMETERS

- A galvanometer is an instrument used to detect the presence of an electric current in a circuit.

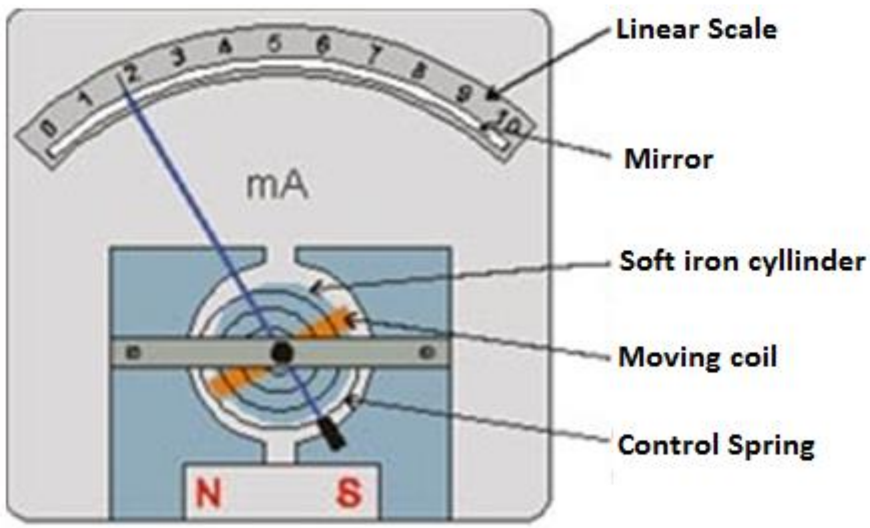
TYPES OF GALVANOMETER

There are two main types of galvanometers. These include;

- Moving iron galvanometer
- Moving coil galvanometer

Moving Iron Galvanometer

- A moving iron galvanometer is a galvanometer where the moving part is iron.



Action

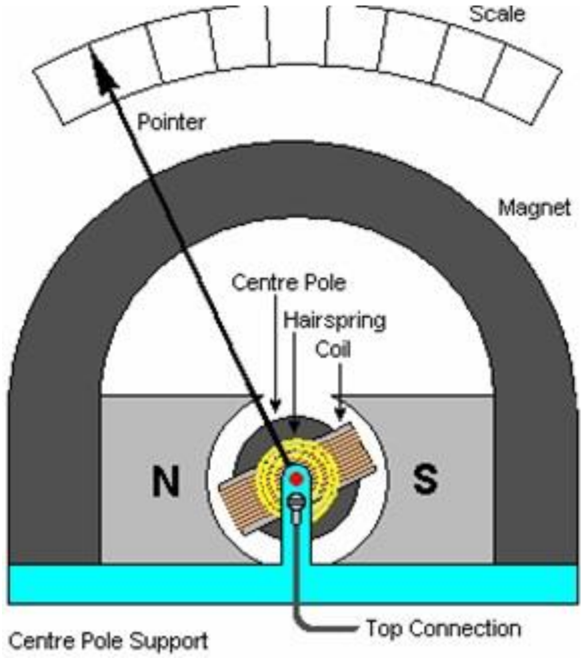
- When the current through the coil it causes the coil to be magnetized hence the movable soft iron will be attracted and the pointer will start moving.

Disadvantages

1. The scale is not uniform.
2. Moving iron galvanometer cannot be converted into ammeter and voltmeter.
3. Part of the energy is wasted in overcoming friction. Hence efficiency is very low.

Moving coil Galvanometer

- A moving coil galvanometer is the galvanometer where by the moving part is the coil.



Action

- The pointer rotates due to the force developed on the coil (by Fleming's left hand rule). When the current passes through the coil will be magnetized hence different poles will be formed.

Advantages

1. The scale is uniform.
2. Moving coil galvanometer is more accurate.
3. The sensitivity of moving coil galvanometer is very high.
4. A moving coil galvanometer can be converted into an ammeter and voltmeter.

FACTORS AFFECTING THE SENSITIVITY OF THE GALVANOMETER

1. The quantity of the current passing through the coil. The bigger is the current the higher is the sensitivity and vice verse.
2. The strength of a magnet. The stronger the magnet, the higher is the sensitivity.
3. The number of turns of the coil. The bigger number of turns in the coil the greater the sensitivity.

4. Cross sectional areas of the coil. The bigger is the cross – sectional area of the coil, the higher is the sensitivity.

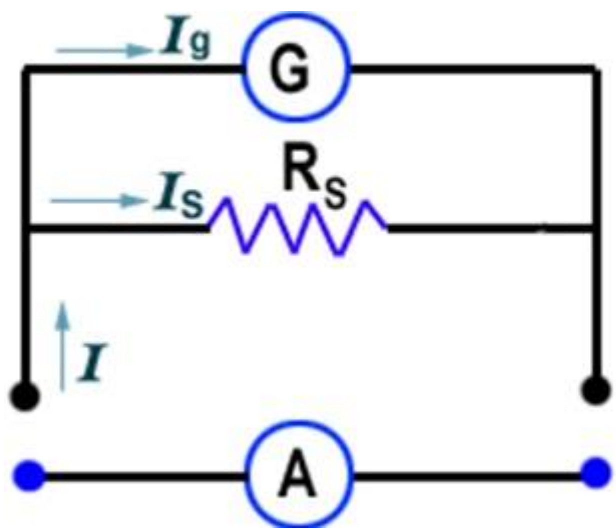
NB: The moving coil galvanometers are multipurpose.

THE CONVERSION OF MOVING COIL GALVANOMETER INTO AN AMMETER

- is an electrical device used to measure the quantity of an electric current in amperes.

HOW

1. The replacement of the wires with low resistance.
2. Connection of the resistor of low resistance called SHUNT parallel to the galvanometer.



If $I = I_S + I_G$

- Potential difference across the shunt is equal to the Potential difference across galvanometer V_{G}.

i.e. $V_S = V_G$

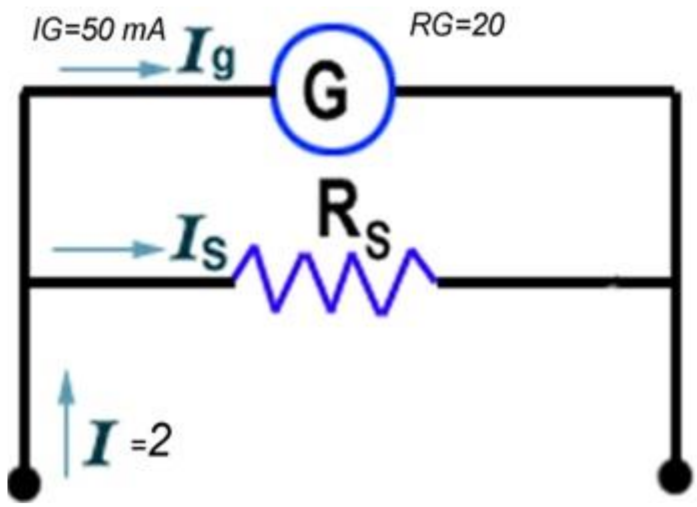
Then

$V_S = I_S R_S$

$V_G = V_G R_G$

Example

1. A galvanometer has a resistance of 50mA passes through it. Calculate the value of the resistance which must be used so that the meter may measures the current up to 2A.



From

$V_S = I_S R_S$

$2 = I_S + 0.05$

$2 - 0.05 = I_S$

$I_S = 1.95A$

$V_G = I_G R_G$

$V_G = 0.05 \times 20$

$V_G = 1.0V$

Ω HENCE : $V_S = V_G$

$V_S = 1.0V$

Therefore

$V_S = I_S R_S$

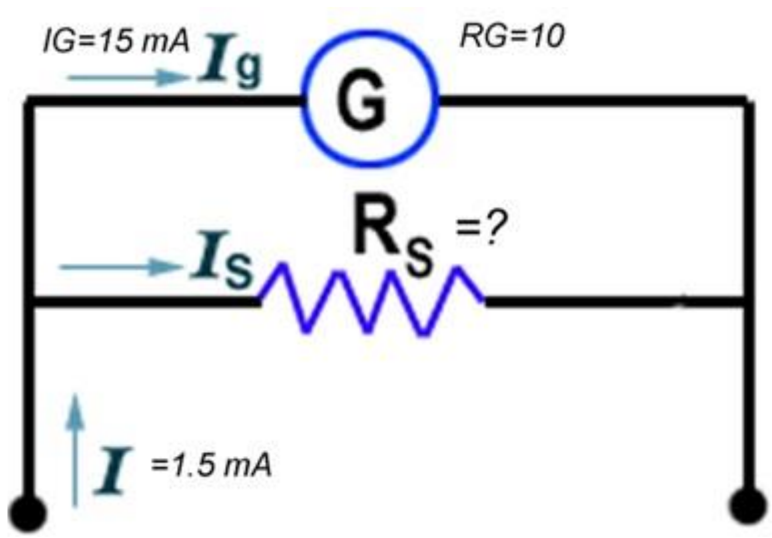
$1 = 1.95 \times R_S$

$$R_S = \frac{1}{1.95}$$

$R_S = 0.5128$

- A shunt of 0.5128Ω must be connected to the galvanometer to give a reading of 2A.

2. A galvanometer coil has a resistance of 40Ω and the full scale deflection current 15mA. If it is to be converted so that it gives a full scale deflection current 1. 5A, then the required shunt will have a resistance of what size?



From; $I = I_S + I_G$

$1.5 = I_S + 0.15$

$I_S = 1.85A$

$VG = I_G R_G$

$= 0.015 \times 10$

$= 0.15V$

From $V_S = V_G$

$V_S = 0.15V$

Therefore

$V_S = I_S R_S$

$0.15 = 1.485 \times R_S$

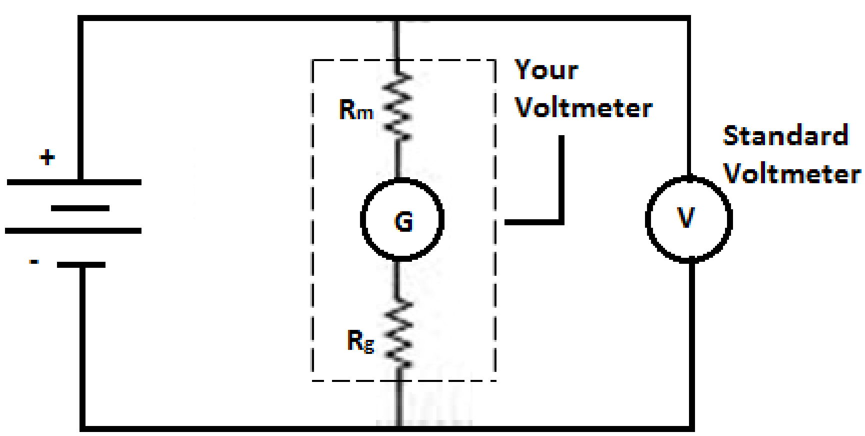
$R_S = 0.1$

THE CONVERSION OF A MOVING COIL GALVANOMETER INTO A VOLTMETER

-A voltmeter is electronically device used to measure the potential difference.

HOW:

1. The replacement of the wire of high resistance.
2. The replacement of a resistor of high resistance known as Multiplier connected in series with the moving coil galvanometer.



3.

Where: R_m = Resistance of multiplier

R_G = Resistance of the coil of galvanometer

V_m = Potential difference across multiplier

V_G = a cross galvanometer

V = Total Potential difference of the full scale deflection

NOTE

$V_m \neq V_G$ (series connection)

But the current, (I) is maintained constant in series circuit

Hence;

$V = V_G + V_m$

and

$V_m = IR_m$

$V_G = IR_G$

Thus;

$V = IR_G + IR_m$

$$\mathbf{V = I(R_m + R_G)}$$

Example

1. A moving will galvanometer has a resistance of 20Ω and gives a full scale deflection when a current of 50mA passes through it.

   Calculate the value of the resistance which must be used so that the meter may measure the potential difference up to 100V.

   $I_G = Im = 50mA = 0.05A$

   From

   $V = I(R_m + R_G)$

   $100 = 0.05(R_m + 20)$

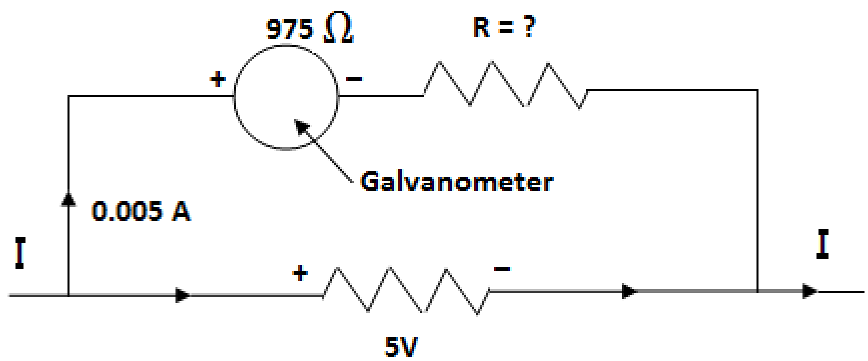
   $100 = 0.05(R_m + 20)$

   $R_m = 2000 - 20$

$R_m = 1980\Omega$

- A multiplier of resistance 1980Ω must be connected in series with the galvanometer so that the meter may measure up to 100V

2. A moving coil galvanometer which gives a full scale deflection of 0.005A is converted to a voltmeter reading up to 5V using an external 975â„¦. Resistance what is the resistance of the galvanometer



Using $V = I(R_m + R_G)$

$5 = 0.005(975 + R_G)$

$1000 = 975 + R_G$

$R_G = 1000 - 975$

$R_G = 25\Omega$

- The resistance of the galvanometer is 25Ω

CLASS WORK

1. A moving coil galvanometer has a coil of resistance 25Ω and can carry a maximum of 15mA.What is the value of the shunt required to enable the galvanometer to register 10A full scale deflection. How will it be connected?

   (b)What is the value of the multiplier required to enable the galvanometer to register 10V full scale deflection. How will be connected?

Solution

1. Shunt Rs =?

$V_G = I_G \times R_G$

$= 0.015 \times 20$

$= 0.3V$

Therefore

$I = I_S + I_G$

$10 = I_S + 0.015$

$I_S = 9.985A$

$V_S = I_S R_S$

$R_S = 0.3 \div 10$

$R_S = 0.03\Omega$

- The value of the shunt required is 0.03Ω. it will be connected in parallel to the galvanometer

b) Using $V = I\,(R_m + R_G)$

$10 = 0.015\,(R_m + 20)$

$666.67 = R_m + 20$

$R_m = 666.67 — 20$

$R_m = R_m = 646.67\Omega$

- The value of the multiplier is 646.67Ω it will be connected in series with the galvanometer.

2. Mention two advantages of alternate current generator over direct current generator

(i) Alternate current generator can be converted to a direct current generator while direct current generator can be converted to alternate current generators because the communications are not complex and expensive.

(ii)The transformers operate on alternate current only for stepping up or stepping down the voltages hence alternate current generator (producing alternative current voltage) is more advantages than a direct current generator (producing a direct current voltage)

RADIOACTIVITY

Atomic theory

In chemistry and physics, the atomic theory explains how our understanding of the atom has changed over time. Atoms were once thought to be the smallest pieces of matter.

The first idea of the atom came from the Greek philosopher Democritus. A lot of the ideas in the modern theory came from John Dalton, a British chemist and physicist.

Democritus' atomic theory

Democritus (Greek philosopher, 460 BC)thought that if you cut something in half again and again, you would at last have to stop. He said that this last piece of matter could not be cut any smaller. Democritus called these small pieces of matter atoms, which means "indivisible". He thought that atoms would last forever, never change and could not be destroyed. Democritus thought that there was nothing between the atoms and that everything around us could be explained if we could understand how atoms worked.

Sir Joseph John Thomson (1856–1940), English physician who discovered the electron and determined its negative charge. He got the Nobel Prize in Physics for 1906.

Dalton's atomic theory

In 1803, the English scientist John Dalton(1766–1844) , reworked Democritus' theory, as follows:

1. All matter is formed of atoms.

2. That atoms are indivisible and invisible particles.

3. That atoms of the same element are of the same type and mass.

4. The atoms that make compounds are present in set proportions.

5. Chemical changes correspond to a reorganisation of the atoms taking part in the chemical reaction.

Dalton defined the atom as the basic unit of an element that can take part in a chemical combination.

In 1850, Sir William Crookes constructed a 'discharge tube', that is a glass tube with the air removed and metallic electrodes at its ends, connected to a high voltage source. When creating a vacuum in the tube, a light discharge can be seen that goes from the cathode (negatively-charged electrode) to the anode (positively-charged electrode). Crookes named the emission 'cathode rays'.

Thomson's atomic model

After the cathode ray experiments, Sir Joseph John Thomson established that the emitted ray was formed by negative charges, because they were attracted by the positive pole. Thomson knew that the atoms were electrically neutral, but he established that, for this to occur, an atom should have the same quantity of negative and positive charges. The negative charges were named electrons (e-).

According to the assumptions established about the atoms neutral charge,

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Thomson proposed the first atomic model, that was described as a positively-charged sphere in which the electrons were inlaid (with negative charges). It is known as the plum pudding model.

In 1906, Robert Millikart determined that the electrons had a Coulomb (C) charge of -1.6 × 10^{19}, something that allowed calculation of its mass, infinitely small, equal to 9.109 ×10^{31} kg.

In the same time, experiments by Eugene Goldstein in 1886 with cathode discharge tubes allowed him to establish that the positive charges had a mass of 1.6726 ×10^{27} kg and an electrical charge of +1.6 × 10 ^{19} C. Lord Ernest Rutherford later named these positively charged particles protons.

Lord Rutherford atomic model

Atomic experiment of Lord Ernest Rutherford

In 1910, the New Zealand physicist Ernest Rutherford suggested that the positive charges of the atom were found mostly in its center, in the nucleus, and the electrons (e-) around it.

Rutherford showed this when he used an alpha radiation source (from helium) to hit the very thin gold sheets, surrounded by a Zinc sulphide lampshade that produced visible light when hit by alpha emissions. This experiment was called the Geiger–Marsden experiment or the Gold Foil Experiment.

By this stage the main elements of the atom were clear, plus the discovery that atoms of an element may occur in isotopes. Isotopes vary in the number of neutrons present in the nucleus. Although this model was well understood, modern physics has developed further, and present-day ideas cannot be made easy to understand.

Modern physics

Atoms are not elementary particles, because they are made of subatomic particles like protons and neutrons. Protons and neutrons are also not elementary particles because they are made up of even smaller particles called quarks joined together by other particles called gluons (because they "glue" the quarks together in the atom). Quarks are elementary because quarks cannot be broken down any further.

Radioactivity is the spontaneous disintegration (break down) of a certain atomic nuclei accompanied with emission of alpha particle (Helium nuclei) beta particle (electron) or gamma rays (electromagnetic wave to short wave length) The atomic nuclei that can undergo spontaneous disintegration are called Radioactive nuclide.

Radioactive decay

Radioactive decay is the spontaneous disintegration of unstable nuclide to form stable nuclide.

TYPES OF RADIOACTIVE DECAY

There are two types of radioactive decay:

1. Natural radioactive decay

2. Artificial radioactive decay

NATURAL RADIOACTIVITY

This is the disintegration of material which occurs in huge unstable nuclide material which emits a particle spontaneous.

The huge unstable nuclides that can undergo natural radioactive decay are:

-Uranium (Ra)

-Radon (Rh)

-Polonium (Po)

-Bismuth (Bi)

-Thulium (Th)

-Actinium (Ac)

The unstable nuclides that can undergo spontaneous decay are those with high mass number. This is because the bond of their nuclide are weakened by the large number of proton forming repulsion force.

EMISSION OF PARTICLE

Radioactive nuclide can undergo spontaneous decay in the form of radiation .These particle are:

- Alpha particle (α – particle)
- Beta particle (β – particle)
- Gamma rays (Ɣ – rays)

1. **EMISSION OF ALPHA PARTICLE**

Alpha particle is denoted by $^{4}_{2}\alpha$ and it is regarded as Helium nucleus ($^{4}_{2}He$)

Thus $^{4}_{2}\alpha = ^{4}_{2}He$

Definition: Alpha particles are helium nuclei emitted by large nucleus during alpha decay. When the nucleus of an atom emit alpha particle it loses 4 unit in its mass number and 2 in its atomic number.

The number left behind will take a different from which can be stable or unstable. If it is unstable it will be also disintegrate until the stable nucleus is obtained.

$$^{A}_{Z}X \longrightarrow ^{A-4}_{Z-2}Y + ^{4}_{2}He$$

Example

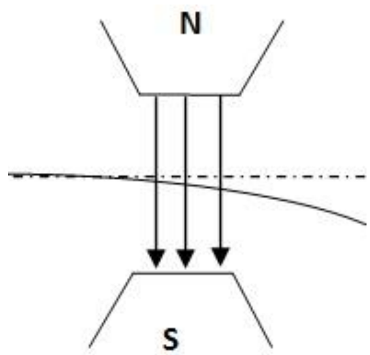
(i) $^{226}_{88}Ra \longrightarrow ^{222}_{86}Rn + ^{4}_{2}He$

(ii) $^{222}_{86}Rn \longrightarrow ^{218}_{84}Bi + ^{4}_{2}He$

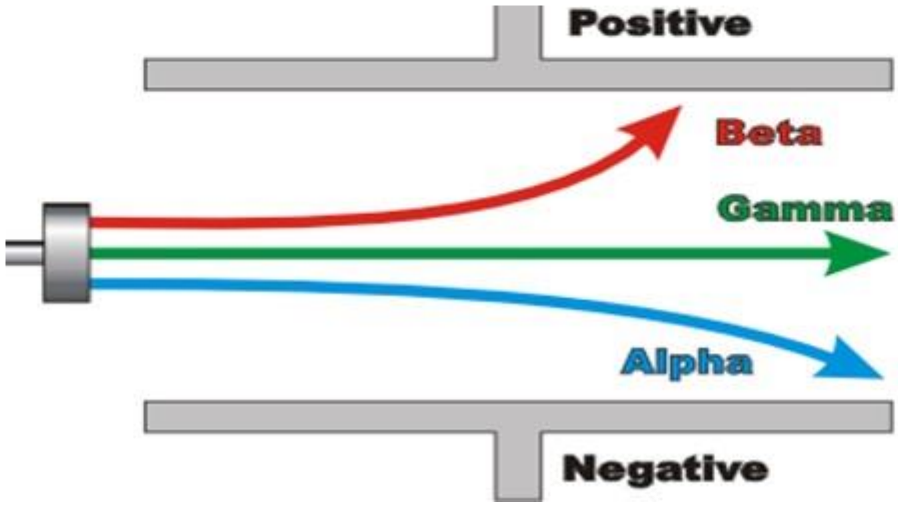
NOTE: A stream of alpha particles is called alpha rays.

PROPERTIES OF ALPHA PARTICLE

1. They are massive particle.
2. The carry positive charge.
3. They can slightly be detected by magnetic field and electric field.



Here α – particle are slightly deflected toward South Pole.



Here α – particle are slightly deflected toward a negative plate.

-The direction of deflection show that α – particle carry positive charge.

4. They cause great deal of ionization in air.

5. They can be absorbed by thin paper.

6. They have low penetration power.

2. **EMISSION OF BETA PARTICLE**

Beta particle is denoted by ${}_{-1}^{0}\beta$ and it be regarded as an electron ${}_{-1}^{0}e$. Thus ${}_{-1}^{0}\beta = {}_{-1}^{0}e$

Beta particle is denoted by ${}_{-1}^{0}\beta$ or ${}_{-1}^{0}e$ and it is regarded as an electron.

Beta particle is an electron emitted by radioactive nucleus during beta decay**.** When the nucleus of an atom emits β – particle the mass number will remain to be the same but the atomic number of the nucleus left behind will increase by 1 unit.

Hence a new nucleus is left behind

$^{A}_{Z}X \longrightarrow {}^{A}_{Z+1}Y + {}^{0}_{-1}e$

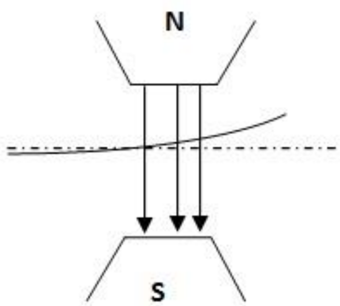
Example

(i) $^{210}_{83}Ba \longrightarrow {}^{210}_{84}Po + {}^{0}_{-1}e$

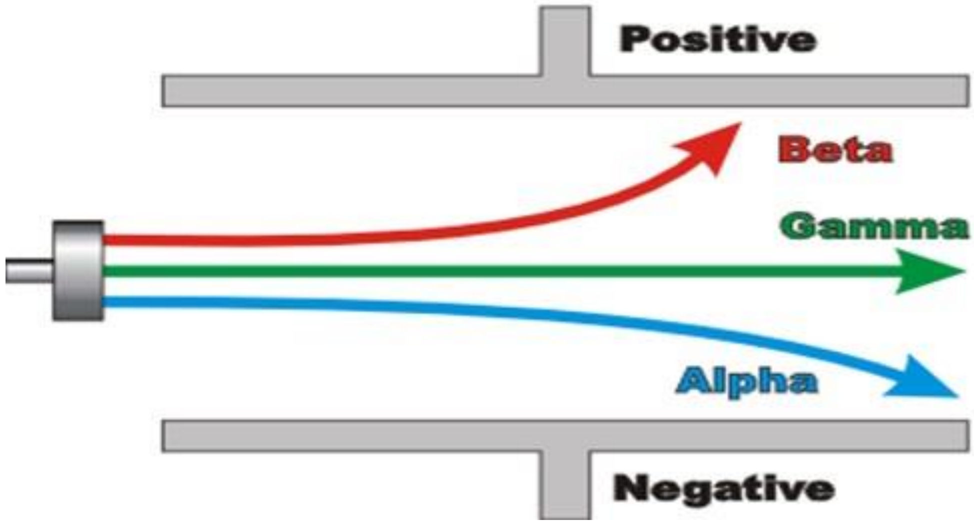
(ii) $^{210}_{82}Pb \longrightarrow {}^{210}_{83}Ba + {}^{0}_{-1}e$

PROPERTIES OF BETA PARTICLE

1. There are mass less particle.
2. They are carry negative charges.
3. They are absorbed by aluminium in few centimeter thick.
4. They can strongly be deflected by magnetic field and electric field.



Here β – particle are strongly deflected toward N – pole.



Here the beam of the β particle is strongly deflected toward a positive plate.

The direction of deflection show that β particle carry negative charge.

5. They cause less ionization in air than α – particle.

6. They have high penetrating power up about 1 meter.

Problem

(a) Uranium $^{238}_{92}U$ decay to polonium by α emission at each stage via $^{234}_{x}Th$, $^{234}_{y}Ra$ and $^{226}_{z}Rn$ .Following this process $^{222}_{84}Po$ decays to $^{q}_{86}Rn$ by emission β– emission only. Write balanced equation of the stage by stage decay process from $^{238}_{92}U$ to $^{q}_{86}Rn$ and hence determine the value of x, y, z, and q

(b) Name isotopes and isobars obtained in the decay process as shown in (a) above

(a) $^{238}_{92}U \underset{\alpha}{\rightarrow} {}^{224}_{84}Po$

$(i)\ ^{238}_{92}U \underset{\alpha}{\rightarrow} {}^{234}_{x}Th + {}^{4}_{2}He$

X + 2=92

X=90

$${}^{234}_{90}Th \underset{\alpha}{\rightarrow} + {}^{230}_{y}Ra + {}^{4}_{2}He$$

y + 2 = 90

y = 88

(iii) ${}^{230}_{88}Ra \xrightarrow{\alpha} {}^{226}_{z}Rn + {}^{4}_{2}He$

Z + 2 = 88

Z = 86

(iv) ${}^{226}_{86}Rn \xrightarrow{\alpha} {}^{222}_{84}Po + {}^{4}_{2}He$

(v) ${}^{222}_{84}Po \xrightarrow{\beta} {}^{q}_{86}Rn + {}^{0}_{-1}e$

${}^{222}_{84}Po \xrightarrow{\alpha} {}^{q}_{84}Rn + + {}^{0}_{-1}e$

q + 0 = 222

q = 222

(b) Isotope ${}^{226}_{86}Rn$ and ${}^{222}_{86}Rn$

Isobars are ${}^{222}_{84}Po$ and ${}^{222}_{84}Rn$

3. **EMISSION OF GAMMA RAYS**

Gamma rays (- rays) are electromagnetic wave of shorter wavelength having the speed of light or are high energy electromagnetic wave emitted by radioactive nucleus.

When the nucleus of an atom emit Ƴ- rays there will be no change in mass number and atomic number of the nucleus Ƴ- rays can never be emitted alone they always come in association with other alpha or beta particle.

PROPERTIES OF GAMMA RAYS

1. They are electromagnetic wave.
2. They carry no charged particle.
3. They cannot be deflected by electric field and magnetic field.
4. They have very high penetrating power and they can only be stopped by thick lead.
5. They cause much less ionization in air than alpha and beta particle.

DETECTION OF RADIATION FROM RADIO ISOTOPES

Radiation from radioactive can be detected by several methods. Some of these methods are:

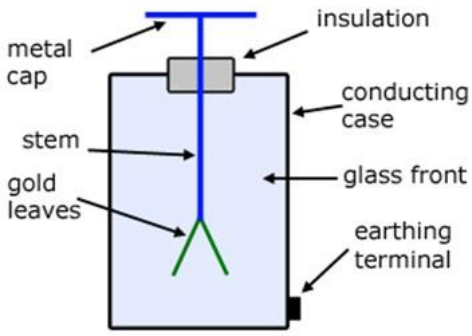
1. Photographic emulsion
2. Gold leaf electroscope
3. Spark counter
4. Geiger Muller tube
5. Diffusion cloud chamber

1. PHOTOGRAPHIC EMULSION

The alpha particle, beta particle, and gamma rays affect the photographic emulsion in a similar way to light.

2. GOLD LEAF ELECTROSCOPE

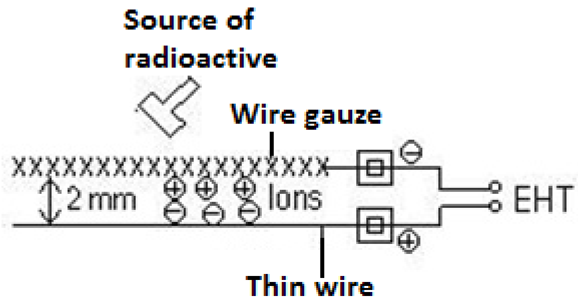
When a radioactive material is brought closer to the metal cap of a charge gold leaf electroscope, the electroscope is slowly discharged this is because the radiation from radioactive material causes ionization of air so that the air becomes a conductor and the charge on the electroscope is emitted through the air.



3. THE SPARK COUNTER

This is an instrument consisting of thin wire a few millimeter away from the plate or is an instrument consist of two parallel electrode 1mm apart.

The wire is kept at high positive potential relative to the plate and almost on the point of sparking.



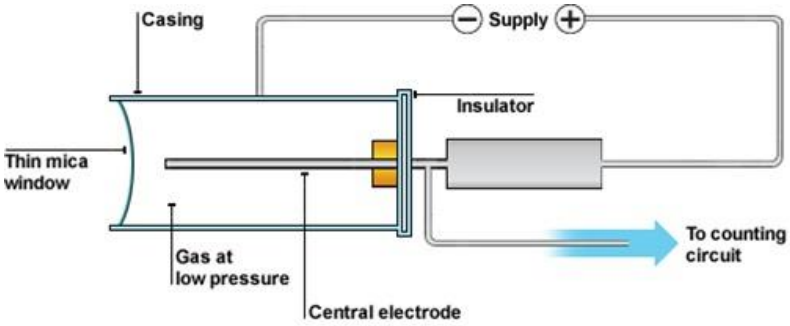
.

-If ionization radiation is passed between the plate and the wire it breaks the insulation of air and spark will be observed.

-The number of spark produced depends on the number of particle produced.

4. **GEIGER MULLER TUBE (GMT)**

The Geiger Muller tube is an instrument which is used to detect the **ionizing properties of radiation**.



When ionization enters a Geiger Muller tube through mica window some argon atom are ionized. The negative ions produced are attracted toward the anode wire and the positive ions are attracted toward the cathode.

A small current in the form of pulse is then produced in the circuit which is amplified and is then sent to the rater meter. The rater meter will count and record the average count rate in count/ sec or counts/min.

Sometimes a small loud speaker is incorporated in the circuits which give pulse for a series.

Back ground radiation

Sometimes Geiger Muller tube gives some background count of radiation even if there is no radioactive material in the neighborhood why?

This is caused partly by radioactive impurities present in the tube and from the surrounding.

5. **Diffusion cloud chamber**

This is an instrument which is used to detect the **individual particles** by providing a record of their track. The instrument consist of glass envelope containing air saturated with mixture of water and ethanol vapor.

The appearance of the cloud tracks in the cloud chamber depends on a particle concerned and it can be used as a mass of identification.

- **FOR ALPHA PARTICLE**

Alpha particle leave straight tracks in a cloud chamber.

Alpha particle

- **FOR BETA PARTICLE**

Beta particle produce wave like tracks in a clouds chamber.

Beta particle

- **FOR GAMMA RAYS**

Gamma rays produce tiny irregular tracks in clouds chamber.

Gamma rays

ARTIFICIAL RADIOACTIVE DECAY

Artificial radioactive decay is the type of disintegration which occur in stable nuclides when stable nuclide are destabilized or is the disintegration which occurs when stable nuclides are destabilized.

When stable nuclides are destabilized they become unstable and they can disintegrate like radioactive nuclide.

Artificial radioactive decay is done by bombarding the nucleus of a stable nuclide by particle such as proton or neutron.

Method / ways of inducing artificial radioactive decay

There are two method / ways of inducing artificial radioactive decay;

1. Bombardment with proton
2. Bombardment with neutron

1. **Bombardment with proton**

Bombardment with proton (^1_1H) is done in a nuclear reactor particle acceleration by a proton (^1_1H) to enter the nucleus of a stable nuclide

Example: $^7_3Li + ^1_1H \longrightarrow ^4_2He + ^4_2He$

Here the nucleus of lithium ^7_3Li undergo disintegration to give α particle ^4_2He

2. **Bombardment with neutron**

Symbol of neutron (n)

This is also done in a nuclear reactor (particle accelerator) by accelerating neutron (^1_0n) to enter the nucleus of a stable nuclide

Example: $^{24}_{12}Mg + ^1_0n \longrightarrow ^{24}_{11}Na + ^1_1H$

Here the sodium isotope is produced and proton is emitted

Now the sodium isotope is $^{24}_{11}Na$ is stable and it can be disintegrated to give back magnesium with ejection of beta particle

$^{24}_{11}Na \xrightarrow{\beta} ^{24}_{12}Mg + ^{0}_{-1}e$

It is more effect to bombard the nucleus of an atom with neutron than with proton.

This is because the bombardment with neutron requires less amount of energy to accelerate neutron to enter the nucleus of a stable nuclide since neutron are neutral in change.

On the other hand bombardment with proton requires large amount of energy in order to overcome the repulsion force between positively charge part of the nucleus and that of the accelerated proton.

HALF LIFE

Half life is time taken by a radioactive material to disintegrate to its half size of a material.

HALF LIFE PERIOD

The half life period of a radioactive sample is that time taken for half the atoms in any given sample of the material to decay.

Each material has its own half life period. Example: The half life of radium is 1600 years while that of bismuth is 10min.

THE HALF LIFE EQUATION

Let $\mathbf{N_O}$ be initial / original number of atoms present in the radioactive sample at time **t = 0**

Let **N** be number of atoms remaining after time t, where t is total time for disintegration

If $\mathbf{T^{1/2}}$ half life of the period of the radioactive sound then

$$N = N_0\left(\frac{1}{2}\right)^{\frac{t}{T^{\frac{1}{2}}}} \text{ ……. ….. half life equation}$$

PROBLEM 1

8×10^8 atoms of Radon were separated from Radium. The half life of Radon is 3.82 days. How many atoms will disintegrate after **7.64 days?**

Data:

Initial number of atom, $N_0 = 8 \times 10^8$

Half life period $\mathbf{T^{1/2}}$= 3.82 days

Total time for disintegration, t = 7.64 days

Solution

Let N be the number of atoms remaining after, t = 7.64days

Let X be number of atoms that will disintegrate after this time

$$X = N_O - N\text{.........eqn (i)}$$

From half life equation

$$N = N_o \left[\frac{1}{2}\right]^{\frac{t}{T\frac{1}{2}}}$$

$N = 8 \times 10^8 \ (1/2)^{7.64/3.82}$

$N = 8 \times 10^8 \times (1/4)$

$N = 2 \times 10^8$ atoms

From equation (i) above

$X = 8 \times 10^8 - 2 \times 10^8$

$= 6 \times 10^8$ atoms

PROBLEM 2

The half life of a radioactive element is 10 minute. Calculate how it takes for 90% of a given mass of the element to decay.

Solution:

Assuming 100% of the element

Initial mass $m_0 = 100kg$

Mass remaining, m = 100kg – 90kg = 10kg

Half life period $T^{1/2}$ =10min

Let t be time taken for 90% of a given mass of the element to decay

From half life equation

$$N = N_0 \left(\frac{1}{2}\right)^{\frac{t}{T\frac{1}{2}}}$$

$N = N_0 (1/2)^{t/T/2}$

$10 = 100 (1/2)^{t/10}$

$10/100 = (1/2)^{t/10}$

$0.1 = 0.5^{t/10}$

Introducing $\log_{10}$ both sides

$$\log 0.1 = \log 0.5^{t/10}$$

$$\log 0.1 = t/10 \log 0.5$$

$$10 \log 0.1 = 1 \log 0.5$$

$$t = \frac{10 \log 0.1}{\log 0.5}$$

Therefore, time = 33min

PROBLEM 3

(a) A radioactive material has a half life of 16 days. How long will it take for the count rate to fall from 160 counts /min to 20counts/min?

(a) Data

Half life period, $T^{1/2}$ = 16days

Initial count rate C_0 = 160counts / min

Final count rate c = 20counts /min

Solution

Let t be the required time

From the half life equation

$$C = C_0\left(\frac{1}{2}\right)^{\left(\frac{t}{T^{1/2}}\right)}$$

$$20 = 160\left(\frac{1}{2}\right)^{\left(\frac{t}{16}\right)}$$

$$\left(\frac{1}{2}\right)^3 = \left(\frac{1}{2}\right)^{\left(\frac{t}{16}\right)}$$

$$3 = \frac{t}{16}$$

t = 3 x 16

t= 48 min

PROBLEM 4:

The half life of the Bismuth is 20min what fraction of a sample of this radioactive bismuth $^{214}_{83}Bi$ remain after 2 hours?

Data

Half life period of Bismuth, $T^{1/2}$ = 20 min

Time for disintegration, t = 2h = 2 x 60 = 120 min

Solution

Let N_0 be initial number of atoms at time t = 0

Let N be number of atoms remaining after time t = 2 hours

$$\text{Fraction remains} = \frac{N}{N_o}$$

From half life equation

$$N = N_0 \left(\frac{1}{2}\right)$$

$$= \left(\frac{1}{2}\right)^{\left(\frac{120}{20}\right)}$$

$$= \left(\frac{1}{2}\right)^{\left(\frac{120}{20}\right)}$$

$$\frac{N}{N_o} = \frac{1}{64}$$

PROBLEM 5

(a) A radioactive nucleus is denoted by the symbol $^{288}_{92}Y$ write down the composition of the nucleus at the end of each of the following stages of disintegration.

(i) The emission of an alpha particle.

(ii) The further emission of a beta particle.

ANSWER

a (i) $\left(^{288}_{92}y \xrightarrow{\alpha} {}^{284}_{90}x + {}^{4}_{2}He \right)$

The composition of the nucleus $^{284}_{90}X$ is 90 proton and $284 - 90 = 194$ neutron

(ii) $^{284}_{90}X \xrightarrow{\beta} {}^{284}_{91}M + {}^{0}_{-1}e$

The composition of the nucleus $^{284}_{90}M$ is 91 proton and $284 - 91 = 193$ neutron

(b) The count rate recorded by Geiger Muller tube and counter close to an alpha particle source is 400 per minute after allowing for the back ground count. If the half life of the source is 4 days.

(i)What will be the count rate 12 days later?

(ii) What should determined over period of several minute rather than over a few second?

Data

Initial count rate C_0 = 400counts/min

Half life of the source $T^{1/2}$ = 4days

Solution

Let C be the count rate after time t = 12

From the half life equation

$$C = C_0\left(\frac{1}{2}\right)^{\left(\frac{t}{T^{1/2}}\right)}$$

$$= 400\ (½)^{12/4}$$

$$= 400 \times \left(\frac{1}{2}\right)^3$$

$$= \frac{1}{8} \times 400$$

C = 50count/min

This is because the rate of emission was so fast.

PROBLEM 6:

A rate meter record a background count rate of 2counts per second when a radioactive source is held near the count rate is 162 counts per second. If the half life of the source is 5minute what will be the recorded count rate be 20min later?

Data

Initial count rate, C_0 = 162 – 2 = 160counts per second

Half life of the source $T^{1/2}$ = 5 min

Total time for disintegration t = 20min

Solution

Let C be final count rate

From the half life equation

$C = 160\ (½)^{20/5}$

$= 160\ (½)^4$

= (1/16) x 160

Therefore C = 10counts /sec

Hence the recorded count rate = 10 + 2 = 12 counts/sec

PROBLEM 7:

A Geiger Muller tube connected to a rate meter is held near a radioactive source. The correct count rate allowing for background count is 400 counts per second. 40 min later the corrected count rate is 25 counter rates per second. What is the half life of the source?

Data

Initial count rate, C_0 = 400couts/sec

Time for disintegration t = 40min

Final count rate C =25counts /sec

Solution

Let $T^{1/2}$ half life of the source

From the half life equation

$$C = C_0\left(\frac{1}{2}\right)^{\left(\frac{t}{T^1/_2}\right)}$$

$$25 = 400\left(\frac{1}{2}\right)^{\left(\frac{40}{T^1/_2}\right)}$$

$$\frac{25}{400} = \left(\frac{1}{2}\right)^{\left(\frac{40}{T^1/_2}\right)}$$

$$\frac{1}{16} = \left(\frac{1}{2}\right)^{\left(\frac{40}{T^1/_2}\right)}$$

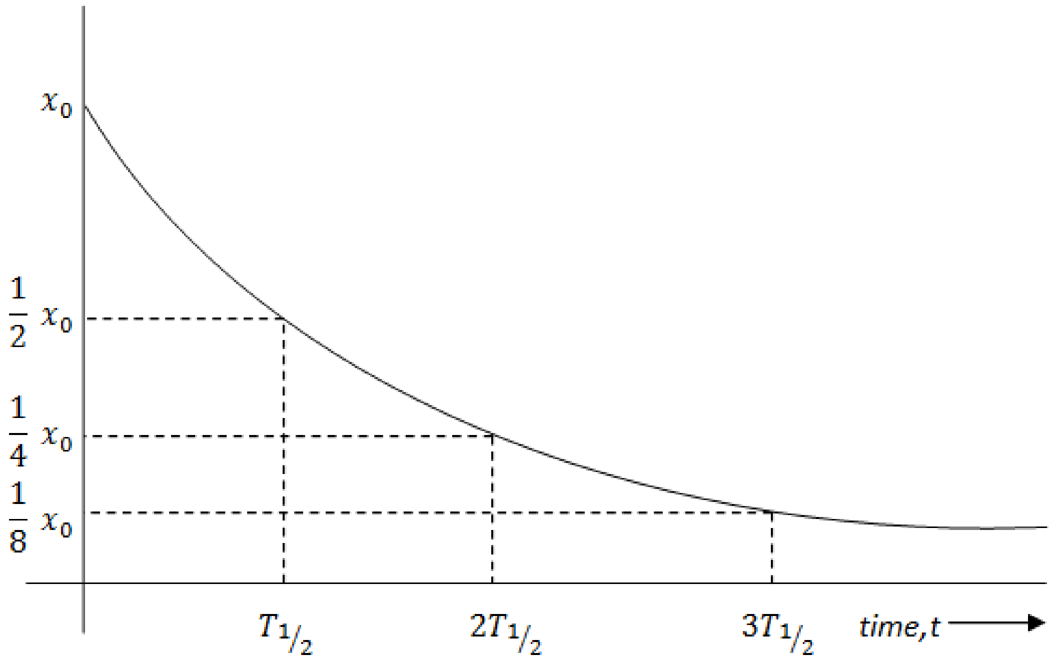
$$4 = \frac{40}{T^1/_2}$$

$$\text{Therefore } T^{1}/_{2} = 10\text{min}$$

THE DECAY CURVE

This is the graph drawn with the number of atoms N present at any time in the vertical axis and the time taken for disintegration in the horizontal axis.

Normally radioactive material never varnishes and hence their graphs with time are asymptotic in nature.



Where by;

$1T_{\frac{1}{2}}$ = First half life period

$2T_{\frac{1}{2}}$ = Second half life periods

$3T_{\frac{1}{2}}$ = Third half life period

NUCLEAR FISSION AND NUCLEAR FUSION

Nuclear fission

Definition:

Nuclear fission is the splitting up of heavy nucleus into two lighter nuclei by neutron capture with emission of neutron followed by energy released.

Example

$${}^{235}_{92}U + {}^{1}_{0}n \longrightarrow {}^{235}_{92}U \longrightarrow {}^{144}_{56}Ba + {}^{90}_{36}Kr + 2\,{}^{1}_{0}n + E$$

Stable nucleus unstable nucleus Lighter nuclei

E = Nuclear energy release

Nuclear energy is that energy released when the nucleus of an atom undergoes disintegration.

Application of nuclear fission

•Nuclear fission is used to produce nuclear energy in nuclear power plant.

Nuclear fusion

- Is the joining (fusing) of two lighter nuclei to form heavy nucleus with emission of neutron followed by energy release.

The fusion of two hydrogen isotopes (Deuterium atom) to give an isotope of helium.

$$\underbrace{{}^{2}_{1}H + {}^{2}_{1}H} \rightarrow \underbrace{{}^{3}_{2}H + {}^{1}_{0}H}$$

Two lighter nuclei   Heavy nucleus

Reason

The speed of approach must be high so as to overcome the repulsion force between positively charge parts of their nuclei.

Nuclear fusion reactions taking place in the interior of the sun produces very large amount of energy which is then used by green plants on the surface of the earth for **photosynthesis.**

SCIENTIFIC APPLICATION OF RADIO ISOTOPES

(1**) In medicine**.

(a) Ƴ – rays from radio isotopes are used in the treatment of cancer by killing cancerous cell.

(b) ϒ – rays from radio isotopes are used to sterilize hospital equipment.

(2) IN INDUSTRIES

(a) Radiation from radio isotope are used to detect the minute cracks or leaks in solid structure.

(b) Radio isotope are used to produce long lasting luminescent paint which can glow in the dark.

(c) Radio isotope are used for the study of wear in machinery.

(d) ϒ – ray from radioactive material are used to control the thickness of paper plastic material and metal sheeting during their manufacture.

(3) IN AGRICULTURE.

(a) Radiation from radio isotopes are used to produce crops with special properties to resist pest.

(b) Radiation from radio isotope are used to examine cracks in a pipe which are used for irrigation purpose.

(4) RADIOACTIVE DATING.

Radio isotopes e.g. carbon fourteen are used to determine the age of ancient material such as rocks, wood etc.

BIOLOGICAL HAZARD OF RADIATION FROM RADIO ISOTOPES

1. They cause diseases which led to death, such as leukemia, cancer etc.
2. Strong doses of radiation from radioisotopes can cause severe burning of the skin and body tissue similar to that caused by fire.
3. They can cause mutation.

SAFETY PRECAUTION

1. Radioactive material are handled by mechanical tongs operated by remote control / equipment with the operator beam being a thick wall of lead or concrete which shields him from the radiation.
2. Radioactive material are stored in thick wall lead container.

THERMIONIC EMISSION

Emission this is the process whereby electrons are emitted (given out) from a substance.

Electron emission this is the process of liberating electrons from the metal surface.

WAYS OF EMITTING ELECTRONS

There are four ways of emitting electrons which are:

- **THERMIONIC EMISSION** Is the process of emitting electrons by applying heat energy. OR is the discharge of electrons from the surfaces of heated materials.
- **PHOTO ELECTRIC EMISSION** Is the process of emitting electrons by application of light energy.
- **HIGH FIELD EMISSION** Is the process of emitting electrons by application of electric field.
- **SECONDARY EMISSION** Is the process of producing electron by application of highest speed field.

GENERAL MEANING OF THERMIONIC EMISSION

At the room temperature, metal consist of electron which can move around a response to an applied electromagnetic field. Under normal condition, the negative charges of the electron are cancelled out by the positive charges in the atom of the metal. If metal is heated the electron gain kinetic energy and can leap out of the metal surface to the surrounding. This phenomena is referred to as thermionic emission

Thermionic emission is the discharge of electrons from the surface of heated materials. This process takes place in a tube called **CATHODE RAY TUBE**

CONDITIONS FOR THERMIONIC EMISSION

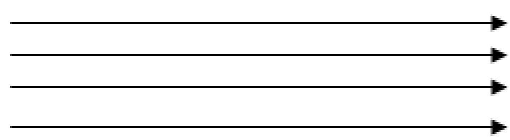
1. The tube must be highly evacuated i.e. low press.

2. The cathode must be hot.

3. There must be Anode which is positive and cathode which is negative.

CATHODE RAY

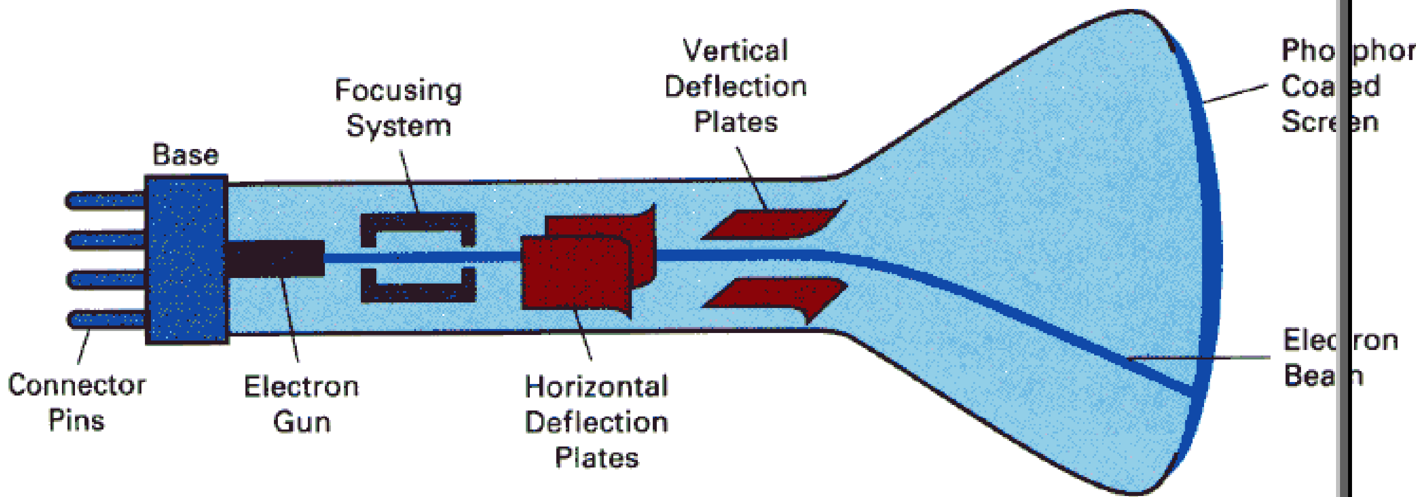
Cathode ray is a stream of fast moving electron. The electron move in a specific direction



Cathode rays

THE CATHODE RAY OSCILLOSCOPE (CRO)

CRO is a substance or tube in which of **cathode rays** are produced. Cathode rays oscilloscope is a **vacuum tube** containing electron gun, deflection system and fluorescent screen. These have internal and external way of accelerate and deflect electron beam that used to form image in form of light emitted and from fluorescent screen.



1. **ELECTRON GUN**

It accelerates and focuses electrons to the screen.

It comprises of;

(a)**Grid** – control the number of electron emitted.

(b)**Cathode** – for electron emitted.

(c)**Anode** – Accelerates and focus electrons to the screen.

The following are the functions of the components of cathode ray tube:

1. **Cathode**

This is a metal filament such as tungsten that is heated to high temperatures either directly by an electric current or directly heating element. The temperature of cathode can range can range from 800^{o}C to several thousands degree Celsius. At these high temperatures some of the valence electrons in the metal attain enough kinetic energy to escape the cathode by thermionic emission. The cathode is maintained at negative voltage.

2. **Anode**

This is a metal disk maintained at a high positive voltage (5000v – 50000v). The anode accelerates the electrons ejected from cathode. There are small opening in the anode through which a narrow beam of electrons passes and enters a region where their direction can be altered.

3. **Horizontal deflection plates**

These are two parallel metal plates carrying equal but opposite charges. They are used to deflect electrons beam horizontally (Left or Right). The beam is attracted to the positive and negative plate and repelled from the negative plate.

4. **Vertical deflection plates**

These are similar to the horizontal deflection plates but oriented to deflect the beam vertically (up or down). The horizontal and vertical deflection plates can direct the beam towards any point on the screen.

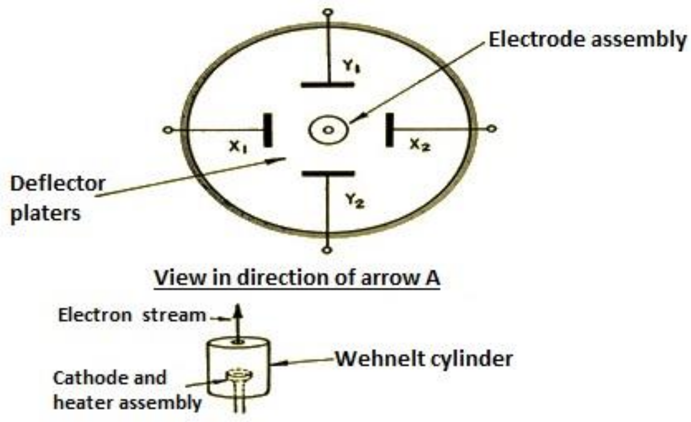
5. **Fluorescent screen**

This is display component of the CRT. It is phosphorus coated so that it emits light wherever the electrons strike it.

The deflection plates move the electrons beam to different points on the screen resulting in the formation of an image .

1. **DEFLECTION SYSTEM**

It controls the deflection of electrons in wrong ways



It comprise of;

(a) Time base it measure the wave forms

(b)Terminal voltage

- Magnetic field like pictures
- Electrical field which are sound

(c)Horizontal plate y - They deflect electron vertically upward and vertically downward

(d)Vertical plate x - They deflect electrons horizontally.

(e)Fluorescent screen

- Gives bright light on the spot
- For displaying signals

$$\frac{1}{2}Mv^2 = ev$$

PROPERTIES OF CATHODE RAY OSCILLOSCOPE

1. Travel in a straight line and they cast shadow.
2. They carry a negative charge.
3. They have energy and momentum.
4. They causes fluorescence (grow) when they strike a materials.
5. They are deflected by electric field and magnetic field.

6. They ionize the gas if potential difference is high and gas pressure is not high.
7. They penetrate in thin sheet of paper or metal foil depend on their energy.
8. They affect photographic film.
9. They produce x ray when stopped suddenly.

Application of the cathode ray tube

The cathode rays tube is used in computer display, cathode television and cathode ray oscilloscopes

1. It can be used as a voltmeter to measure voltage
2. Display waveform
3. to measure time intervals
4. To measure phase relationship
5. Comparison of frequencies

Televisions and Computer Monitors

In television sets and computer monitors, the entire front area of the tube is scanned repetitively and systematically in a fixed pattern called a raster. An image is produced by controlling the intensity of each of the three electron beams, one for each additive primary color (red, green, and blue) with a video signal as a reference. In all modern Cathode Ray Tube(CRT) monitors and televisions, the beams are bent by magnetic deflection, which is a varying magnetic field generated by coils and driven by electronic circuits around the neck of the tube

Monochrome Computer CRT Monitor

Monochrome monitor - this CRT uses only one type of phosphor. Although a mainstay of display technology for decades, CRT-based computer monitors and televisions constitute a dead technology. The demand for CRT screens has dropped precipitously since 2000, and this fall off has been accelerating in the latter half of that decade. The rapid advances and falling prices of LCD flat panel technology, first for computer monitors and then for televisions, has been the key factor in the demise of competing display technologies such as CRT, rear-projection, and plasma display.

Oscilloscope

An oscilloscope is a device t**hat measures** and **displays voltages** as a time versus voltage graph. The voltage difference between the positive and negative probe leads is measured, buffered, and displayed on the screen as a continuous curve. Oscilloscopes are generally used to see if a circuit is performing as expected, but oscilloscopes are also useful for comparing different signals to each other

Oscilloscope Display

Example of an analog oscilloscope display Shown is a Lissajous figure, showing a harmonic relationship of one horizontal oscillation cycle to three vertical oscillation cycles.

Many oscilloscopes also use CRT displays, though LCD displays are becoming more common. In oscilloscope CRTs, electrostatic deflection is used, rather than the magnetic deflection commonly used with television and other large CRTs. The beam is deflected horizontally by applying an electric field between a pair of plates to its left and right, and vertically by applying an electric field to plates above and below.

Oscilloscopes uses electrostatic rather than magnetic deflection because the inductive reactant of the magnetic coils would limit the frequency response of the instrument. The color of the oscilloscope phosphor is much less important than in the case of color televisions or computer monitors since the primary purpose is to evaluate signal voltages rather than construct complex images; however, the persistence of the phosphor may be more important. Phosphors are available with persistence ranging from less than one microsecond to several seconds. For visual observation of brief transient events, a long persistence phosphor may be desirable. For events which are fast and repetitive, or high frequency, a short-persistence phosphor is generally preferable.

X – RAYS

X-radiation (composed of X-rays) is a form of electromagnetic radiation. Most X-rays have a wavelength in the range of 0.01 to 10 nanometers, corresponding to frequencies in the range 30 petahertz to 30 exahertz (3×1016 Hz to 3×1019 Hz) and energies in the range 100 eV to 100 keV. X-ray wavelengths are shorter than those of UV rays and typically longer than those of gamma rays.

X – Rays are electromagnetic radiation which produced when cathode ray stopped rapidly by hard object. X – Rays are reflected rays when cathode ray hits (falls) the metal target.

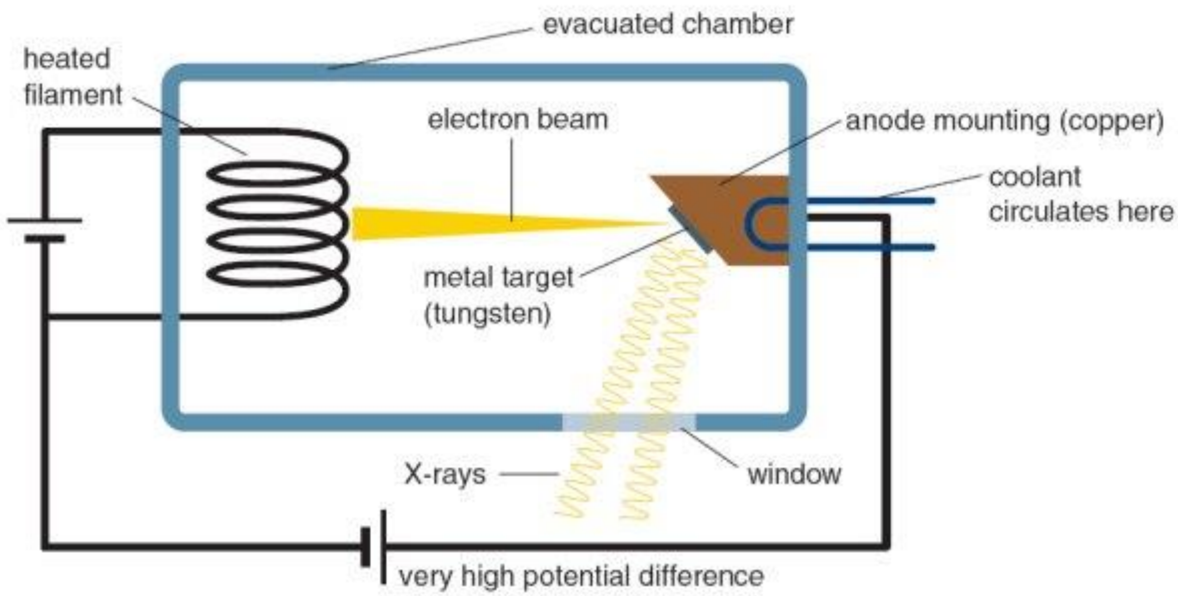
THE X – RAYS RESULT FROM TWO PROCESS NAMELY

1. The rapid slowing down of electron as they enter the target Atom.

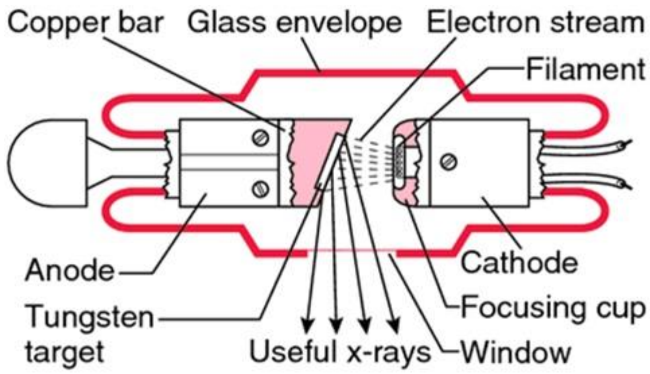
2. The excitation of the target Atom.

X – Rays is a reflected rays when the cathode rays hits (falls) the metal target. X – Rays are produced x – rays tube.

CONSIDER THE DIAGRAM BELOW OF THE X – RAYS TUBE BELOW



Alternative diagram



X – RAYS TUBE CONSIST OF;

1. Heater – produce heat.

2. Glass tube – evacuated glass tube to keep out gas molecules.

3. Concave cathode – focusing cathode rays (electron) to a spot on tangent.

4. Cooling firm – To remove much of the heat conducted along the thick copper rod.

5. Tungsten target – A target in which its atom when strike by electrons excited after absorbing K.E and converted into x – rays radiation.

6. Copper rod – conduct heat away from the target.

HOW X – RAYS TUBE ARE PRODUCED

X-rays are produced when electrons beam strike a metal target. The electrons are liberated from the heated filament and accelerated by a high voltage towards the metal target. The X-rays are produced when the electrons collide with the atoms and nuclei of the metal target.

The anode is maintained at high potential so as electrons accelerates at the speed necessary to produce the x – rays. Only small fraction of K.E of the electrons becomes X – Rays radiation the rest is absorbed by the target which becomes hot.

PROPERTIES OF X – RAYS

1. They penetrate through substance but absorbed more by dense solid.
2. Affect photographic film.
3. Ionize gases (so that the gases become conductor).
4. Not deflected by magnetic or electric field.
5. Carry no charges (neutral).
6. The speed of X – rays is 3 x 10^8m/s to give a more intense beam of x – rays the cathode must be made hotter to give more electrons and give more x – rays.

TYPES OF X – RAYS

X – Rays have wavelengths between 10^{-8}M to 10^{-10}M within this range we can get two types of X – rays due to different in wavelength and frequency these are:

1. hard x ray and
2. soft x ray

1. **SOFT X – RAYS**

This is the one which have long wavelength but lower range of frequency. Soft x ray produced by lower voltage and has less penetrating power.

- Have longer wavelength (10^{-8}M) e.g. TV set emit small amount of soft rays.

2. **HARD X – RAYS**

This is the one which have got short wavelength high range of a frequency. Produced with high voltage and has high penetrating power.

- Have short wavelength (10^{-10}M).

Uses of x – rays

x – Rays finds many applications in hospitals industries even in scientific research:

1. **Medical practice**

- To detect broken teeth or bones
- With case x – rays can be used to kill cancer cells and tumour cells

2. **In industries**

- To detect the broken part of a machines
- x – rays machines are used to reveal hidden metal flaws
- x – rays are used to reveal defects in steel plates

3. **In scientific research**

- x – rays microscopes have made it possible to study the arrangement of possible to study the arrangement of the molecules of crystalline substance

4. **In agricultural activities**

5. **In science and technology**

HAZARDS/ EFFECTS OF X – RAYS

The part which get x – rays the body tissues are destroyed and kills cells which cause the cancer cell one hard x – rays reduce 3 yrs of living.

X – Rays are dangerous to us because use our bodies absorb the energy from x – rays radiations when bodies absorb the x – rays energy ions are produced in the body. These ions can change or destroy living cell. The damage to the body's living cell can stop them from functioning and multiplying which can lead to;

1. Cancer

2. Leukemia (blood cancer)

3. Hereditary defects in children

4. Death

PRECAUTIONS

People are therefore advised against exposing themselves to X – Rays unless it is absolutely unavoidable.

BASIC ELECTRONICS

Electronics is a branch of physics that deals with the emission and effects of electrons in materials.

Or

Is a branch of science dealing with the study and development of circuit involving semi conductors, logic gates and other electrical components like resistors, capacitors and inductors.

Or

This is the branch of physics which deals with the movement of electricity in different materials.

Electronic system or circuit is made up various components connected to each other. They are used to perform a wide variety of tasks. The main uses of electronic circuit are;

1. Conversion and distribution of data.

2. Controlling and processing of data.

Electronic components can be passive or active:

1. **PASSIVE COMPONENTS**

   Consume but do not produce energy. They do not have the ability to produce GAIN that is to increase the power or AMPLIFY the signal. They also do not have directionally, that's they operate in the same way regardless of the direction of the current flowing through them passive components include power sources (battery or generator), resistor, capacitor and inductors.

2. **ACTIVE COMPONENTS**

These are those that have direction and or the capacity to produce gain. They include semi conductor devices such as diodes, transistors and integrated circuits.

Additional terms

Voltmeter – is the device that measure potential difference.

Ammeter – is the device which record electric current.

Inductors – These are coils wound out soft iron core used to control alternating current by self induction.
Also used to minimize effect of excessive alternating current in the electric circuit. Usually
inductors are made up by a copper wire.

ENERGY BANDS IN SOLIDS

The energy band in sold is the range of energy possessed by an electron in solid materials (crystals).

This happens when atoms of element combine to form solid materials, they arrange themselves in an orderly manner called crystal. Therefore the energy levels of the electrons of the atoms in a solid are modified and thus each electron in any orbit of an atom can have a number of discrete but closed spaced energy level lying within a certain range.

The energy bands formed when the atom combined to form a solid one of three categories;

1. Valence band
2. Conduction band
3. Forbidden gap

An important parameter in the band theory is the Fermi level, the top of the available electron energy levels at low temperatures. The position of the Fermi level with the relation to the conduction band is a crucial factor in determining electrical properties.

1. **VALENCE BAND**

This is the range of energy possessed by valence electron. It has the electrons of highest energy. The band may be completely or partially filled with electrons, where the electrons are normally present at the absolute zero temperature.

2. **CONDUCTION BAND**

This is the range of energies possessed by conduction free electrons. The band may be partially or not filled with electrons.

3. **FORBIDDEN ENERGY GAP**

This is the energy of separation between conduction band and valence band.

No electrons of solid can stay in this band.

CLASSIFICATION OF SOLIDS WITH ENERGY BANDS

Due to the arrangement of electrons in energy bands leads to the formation of three types of solids in terms of conductivity which are;

1. Conductors (metal)
2. Semi conductors

3. Non – conductors (or insulators)

1. **ENERGY BAND IN CONDUCTORS (METAL)**

CONDUCTORS

This is the material which can conduct electricity at any temperature.

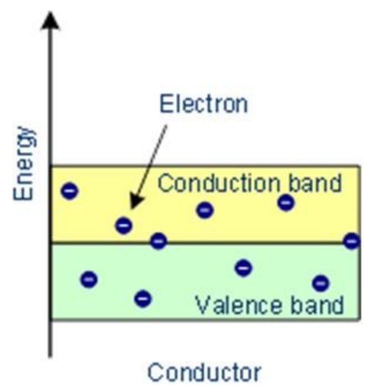
Or

Conductors are those substances which easily allow the passage of current by means of free electrons through them.

- Absence of forbidden gap.
- Valence band and conduction band overlap one another.

PROPERTIES OF A CONDUCTOR

- Its electricity conductivity decrease with temperature.
- This is due to some electrons loss amount of energy due to collision.



2.ENERGY BAND IN THE NON – CONDUCTOR (INSULATED)

NON – CONDUCTOR

This is the material which cannot conduct electricity at any temperature.

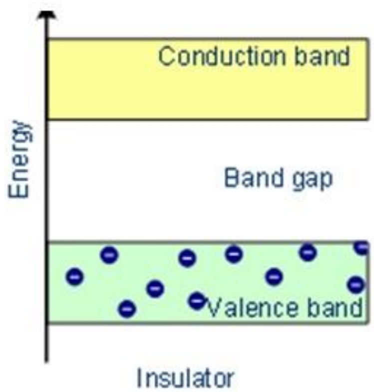
Or

These are the materials which do not allow the passage of electric current by means of free electrons through them.

- The energy band in non – conductor arranged as follows.
- The valence band is completely filled with electrons.
- The conduction band is empty.
- The forbidden energy gape is large about 15ev. And therefore valence electrons from the valence band can never gain enough energy to overcome the forbidden energy gap.

PROPERTIES OF INSULATOR

1. The forbidden gap is higher which makes a given electrons an efficient to jump from valence band to conduction band.
2. Presence of higher electrons at in force.
3. Cannot conduct electricity.



ENERGY BAND IN SEMICONDUCTOR

SEMICONDUCTOR

This is a material which behaves as an insulator at OK and conductor at 273K.

Or

This is the material which has the properties lies between that of insulator and conductor.Examples of semiconductors are silicon, germanium, cadmium sulphide and gallium arsenide.

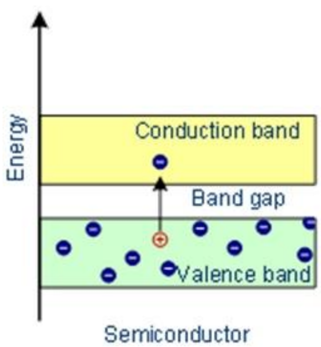
a) At absolute zero temperature (OK)

A semi conductor acts like a non – conductor. The valence band completely filled with electrons, and conduction band is completely empty. Forbidden gap is wide.

b) At room temperature

A semiconductor acts like a conductor:

- The valence band is partially filled.
- Conduction band has few electrons.
- For bidden energy gap is narrow.



PROPERTIES OF SEMICONDUCTOR

1. Presence of narrow for bidden gap.
2. Its conductivity increases with temperature.
3. Presence of partial electrons in the conduction on band.
4. They have negative temperature coefficient of resistance.

EFFECTS OF TEMPERATURE IN SOLIDS

1. **EFFECTS OF TEMPERATURE IN NON CONDUCTOR (INSULATED)**

The temperature has no effect on the conductivity property of the insulator since the forbidden energy gap is very large.

2. **EFFECTS OF TEMPERATURE IN A CONDUCTOR**

The conductivity of the conductor decrease as the temperature increases. Since when the temperature increase, it rises the amplitude of vibration of atoms and more collision with atoms are made by drifting electrons and this slow the free electrons and hence conductivity decrease.

3. **EFFECTS OF TEMPERATURE IN SEMI CONDUCTORS**

At absolute zero temperature.

At this temperature the valence band is full filled and there is a large energy gap between valence and conduction's band. There is no valence electron can reach the conduction band to become free electron, thus the material behaves like an insulator.

Above absolute zero temperature, as the temperature rises some valence electrons acquire sufficient energy to enter into the conduction band and this becomes free elections. Thus the conductivity increases as the temperature increases.

TYPES OF SEMICONDUCTORS
There two types of semiconductors:-

a) **Intrinsic semiconductor**

These are pure semiconductors which the charge carrier originates inside the material itself. The conduction of electricity takes place by the promotion of electrons from the valence to the conduction band energy bands in intrinsic semiconductor. The conductivity in this semiconductor is due to increase in temperature and the main charger carrier is electron.

In intrinsic semiconductor, the number of free electrons and the number of holes are exactly equal.

b) **Extrinsic semiconductors (Impure semi conductor)**

This is type of a semiconductor which is conduction are introduced and improved through the doping process. In these types of semiconductor you can modify conduction level by adding small amount of impurity up to a million times.

DOPING

Doping is process of adding impurity atoms to intrinsic crystal to produce an extrinsic semiconductor.

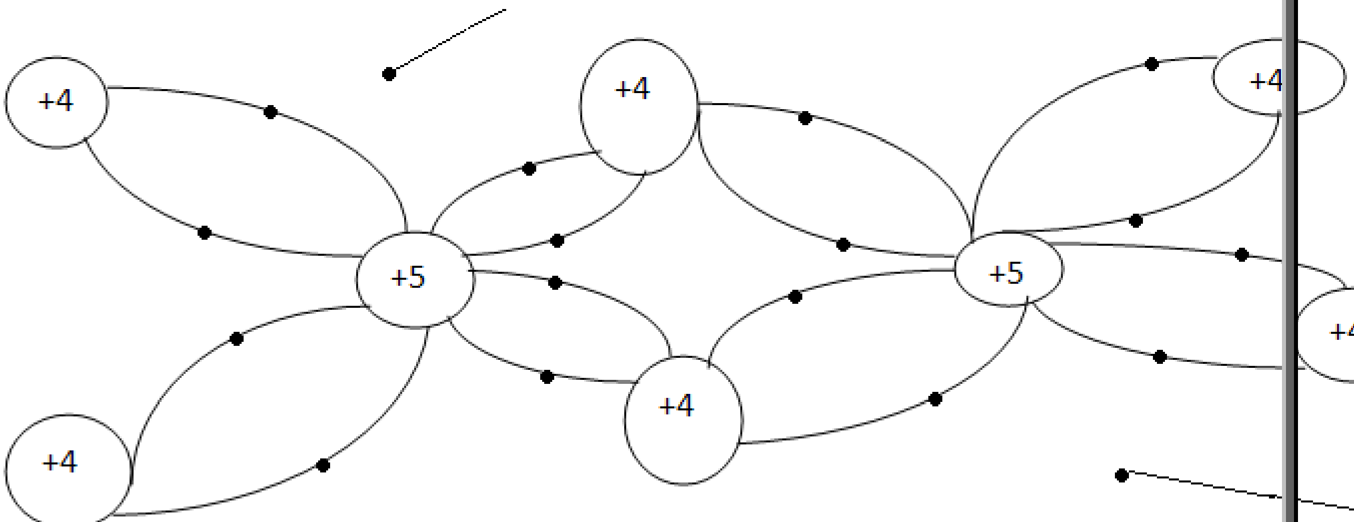
OR

This is the process of adding an impurity is to increase the number of free electrons or holes in the semiconductor. The purpose of adding impurities is to increase the number of free electrons or holes in the semiconductor in order to increase the conductivity of the semiconductor. The impurities are called dopants.

MECHANISM OF DOPING SEMICONDUCTORS

Both silicon and germanium are tetravalent atoms, ie they have for valence electrons (four electrons to the outermost shell of the atom). If pentavalent atom (eg. phosphorus) replace a silicon semiconductor, four of the impurity electron play the same role as the four valence electrons of the replaced sililcon atom and become part of the valence band. The fifth valence electron is easily detected from the pentavalent atom by thermal energy and moves freely in the conduction band. Impurities that denote electrons to the conduction band are called donor impurities.

Free electron



Free electron

Fig 1.0

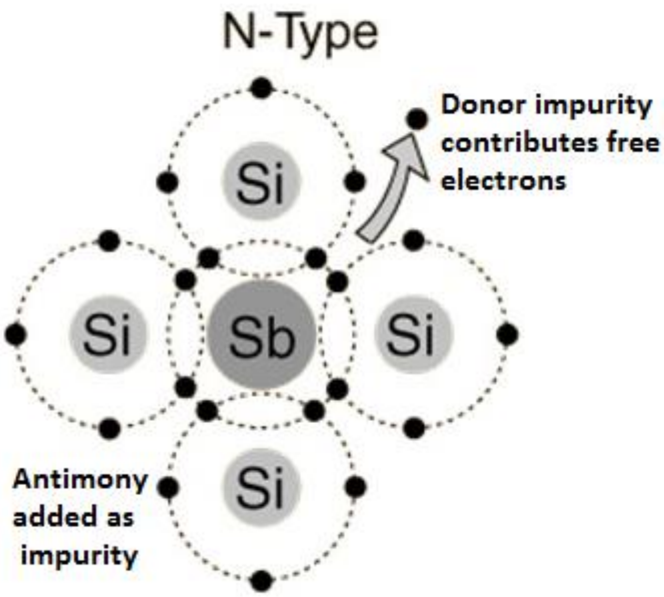
Effect of adding a donor impurity to a silcon semiconductor.

Doping produces two types of semiconductors which are:-

1. n – type semiconductor
2. p – type semiconductor

This is the semiconductor which the majority charge carriers are electrons. (Negativity charges)

The **n – type semiconductor** obtained by adding the pentavalent element to the pure semiconductor (Trivalent element) the addition of pentavalent impurities provides a large number of free electrons in a semi conductor. The pentavalent impurities are called **donor**s since they provide free electrons to the semiconductors. Suppose a pentavalent element e.g. antimony atom is added to a pure germanium trivalent atom s shown below;



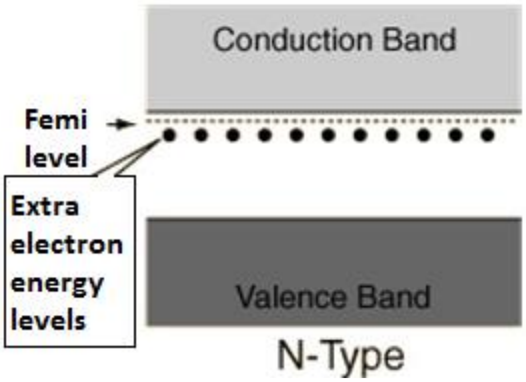
Si – silicon

Sb – Antimony atom

The form valence electrons of Antimony will form bonding with the germanium valence electrons. The fifth valence electron of antimony is not involved in bonding, so remain free to move thus the number of electrons carries increases and hence the conductivity of material increases.

ENERGY BAND OF N – TYPE SEMICONDUCTOR

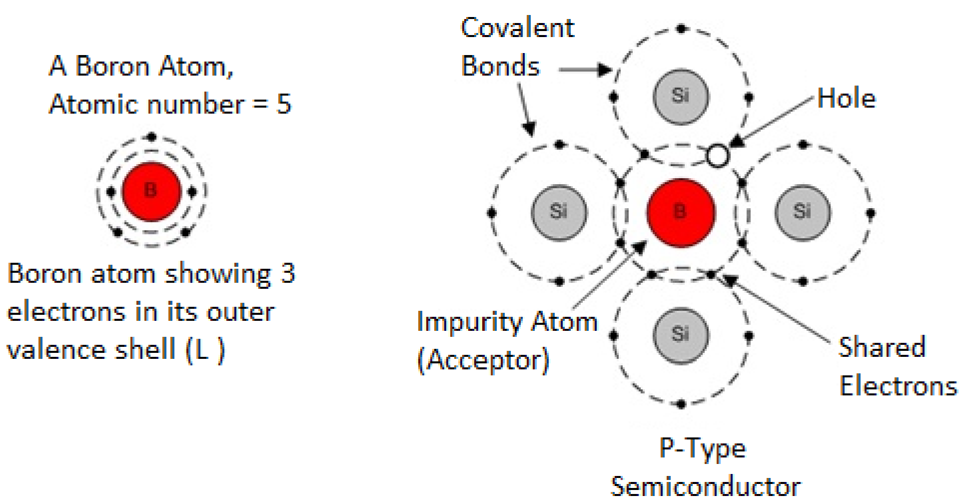
The additional of donor impurities to an intrinsic semiconductor, creates extra energy level called donor energy level just below the bottom of the conduction band at the forbidden band.



1. **P – Type of semiconductor**

This is a semiconductor which the majority charge carries are positive charges holes. The **p – type semiconductor** obtained by adding trivalent impurities to the pure semiconductor. The additional of trivalent impurities provides a large number of holes in the semiconductor.

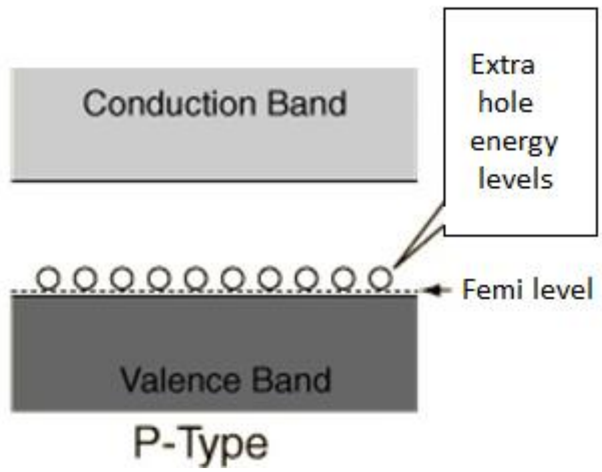
The trivalent impurities are called **an acceptor** since it receives the electron from the semiconductor. Suppose a trivalent element e.g. indium atom is added to a pure semiconductor of germanium (trivalent) atom as shown below;



The three valence electron of germanium form complete bonding in the indium the fourth bond is complete being short of one electron. This missing electron is called **a hole**. Therefore each indium atom added one hole is created. The number of positive charge carrier's increases and this increases the conductivity of the material. The charge carries of current are positive charge.

ENERGY BAND IN P – TYPE SEMICONDUCTOR

The additional of acceptor impurity to an intrinsic semiconductor creates extra energy level called **acceptor energy level** just above the top of the valence band



COMPARISON BETWEEN EXTRINSIC AND INTRINSIC SEMICONDUCTOR

INTRINSIC SEMICONDUCTOR	EXTRINSIC SEMICONDUCTOR
It is a pure semiconductor.	It is an impure semiconductor.
The number of electricity equals to the number of holes.	The number of free electrons not equal to the number of holes.
The electric conductivity is low.	The electric conductivity is high.
The electric conductivity depend on temperature.	The electric conductivity depend on the temperature and amount of doping.
It has no practical use.	Used in electronic device.

COMPARISON BETWEEN N – TYPE AND p – TYPE SEMICONDUCTOR

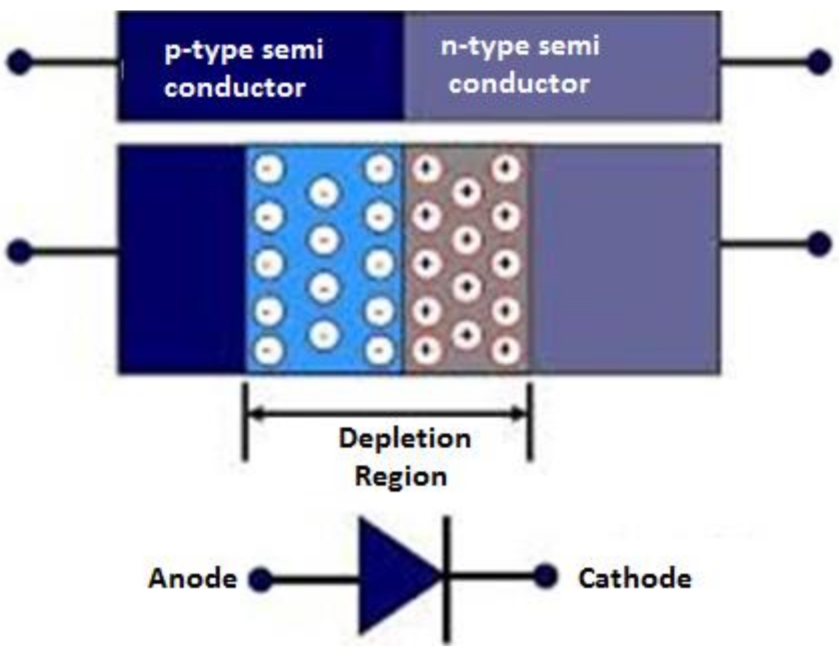
N – TYPE	P - TYPE
Produced by adding pentavalent impurities to a pure semiconductor.	Produced by adding trivalent impurities to a pure semiconductor.
The number of free electron exceed the number of holes.	The number of holes exceeds the number of free electrons.
The majority charge are negative charges.	The number of holes exceed the number of free electrons.
The donor energy level is just below the bottom of the conduction band.	The acceptor energy level is just above the valence band.

1. JUNCTION DIODE

This is the p – n junction semiconductor material which is connected to supply voltage.

P–N junction

This is the junction made up by two semiconductor material of n – type and p – type melted together to form a junction.



As soon as a p – n junction is formed electrons from n – type materials diffuse into p – type material and fill some of the holes there At the same time holes from p – type materials diffuse into n – type materials and are filled by electrons This diffusion establishes a potential differences across the junction and within a very short time of the junction being made this become a large enough to prevent any further movement of charge carriers this p–d is called **barrier** or **junction barrier**.

DEPLETION LAYER

This is the region of the P-N junction at which charges exchange direction of moment. The potential difference set up at the junction is called **POTENTIAL BARRIE**.

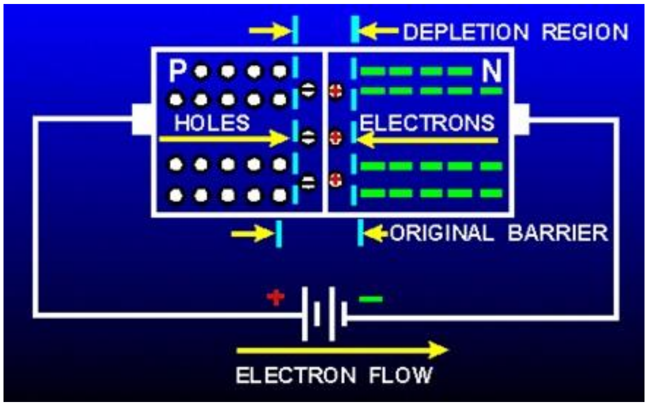
BIASING OF A P – N JUNCTION

This is the circuiting process of a semiconductor device such that it can either highly allow or highly prevent movement of charges through it. Is the process of applying potential to the pn–junction there are two types basing which are;

1. Forward biasing
2. Reverse biasing

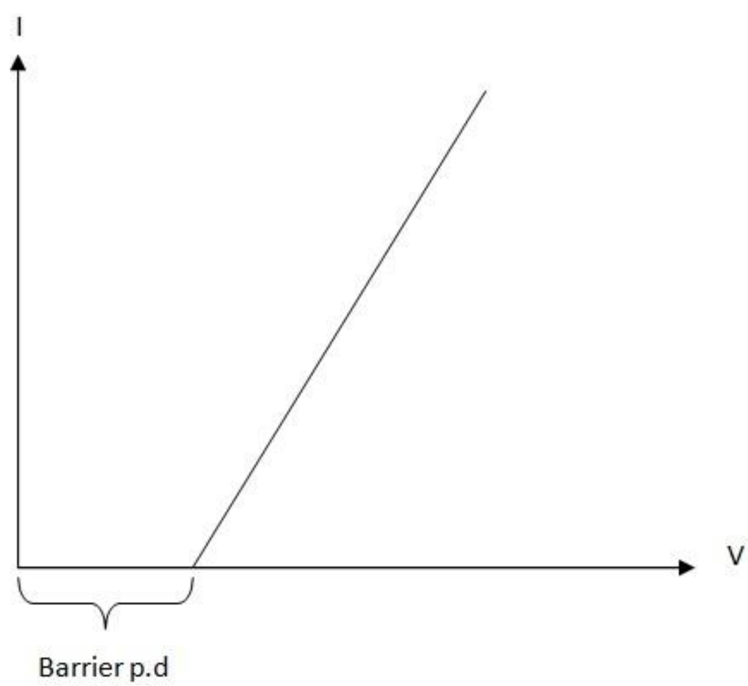
1. FORWARD BIASING

This is the process of connected the positive terminal of the battery to the p–type end and negative terminal of the battery to the n–type end of the p-n junction.

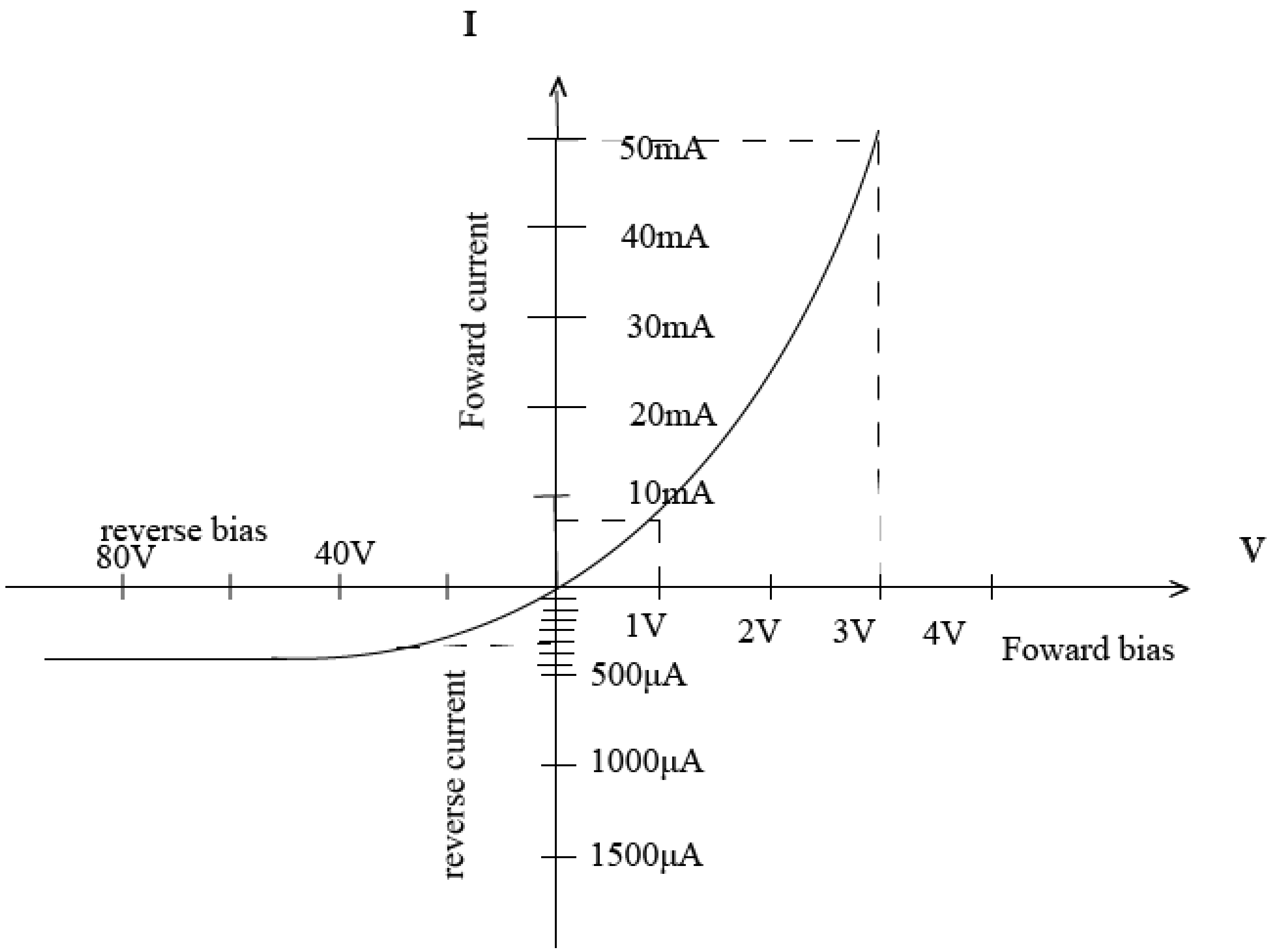


FLOW OF CURRENT IN A FORWARD BIASED P-N JUNCTION

Under the influence of forward voltage the free electrons in n–type move forward the junction and holes in p-type move forward the junction and due to the large number of concentration of the charge, the free electrons and holes cross the junction and constitute the current. Thus in n–type region current is carried by free electrons and in p–type region the current is carried by holes. Thus the forward biased pn-junction allows the current to pass through the junction.



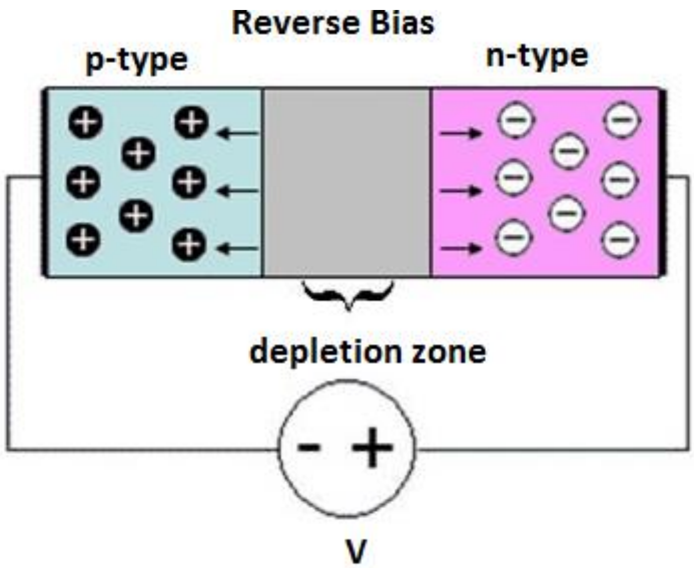
Forward biased



Forward and Reverse bias

1. **REVERSE BIASING**

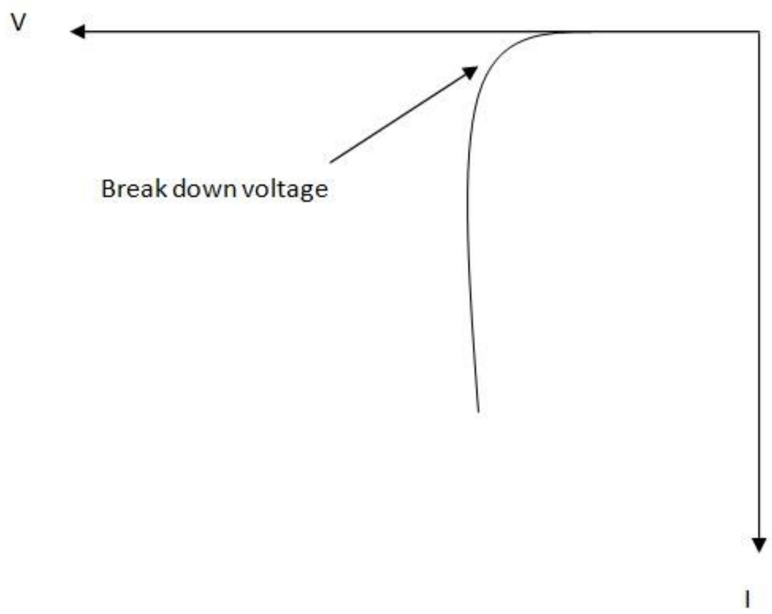
This is the process of connecting the positive terminal of the battery to the n–type and negative terminal of the battery to the p–type of the p-n junction.



FLOW OF CURRENT IN THE REVERSE BIASED P – N JUNCTION

With reverse bias to the p-n junction, the potential barrier at the junction is increased and practically no current flow through the circuit by majority charge carrier. However in practice a very small current flow in the circuit due to the minority carrier. If the reverse voltage is increased continuously the ice of minority electrons may become high enough to knockout electrons from the semiconductor atom.

AI – This stage break down of the junction occurs characterized by a sudden rise of reverse current and a sudden fall of the resistance of barrier region. This may destroy the junction permanently.



- **Break down voltage**

This is the reverse voltage at which p-n junction break down with the sudden rise in reverse current.

- **Knee voltage**

This is the forward biased voltage at which the current through the junction starts to increase rapidly.

DIODES

A diode is an electrical device allowing current to move through it in one direction with far-greater ease than in the other. The most common kind of diode in modern circuit design is the semiconductor diode, although other diode technologies exist. The diode they work on the p-n junction work. The n-region is called the cathode and the P-regions is the anode.

Block diagram

PN diode and its symbol

Note:
(+) Anode and (-) Cathode.
N-Region represent Cathode and P-Region represent Anode.

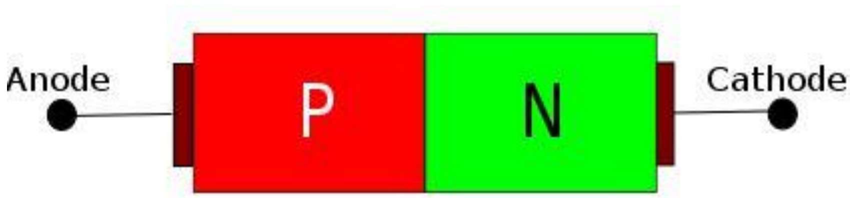
TYPES OF DIODE

There are different types of diodes used in the electric circuits. The following are the most common ones:-

1. Semiconductor diode
2. Metal semiconductor diode
3. Zener diode
4. Light emitting diodes

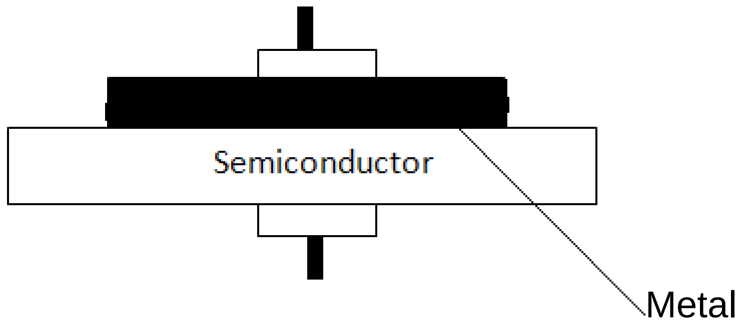
1. **Semiconductor diode**

One type of diode of diode is p-n junction diode described above. This is referred to as semiconductor diode . Most semiconductor diode are made up of silicon or germanium.Consider the figure below shows a construction of Semiconductor diode.



2. **Metal semiconductor diode**

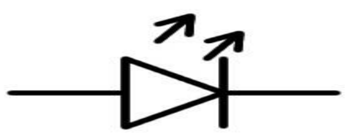
These types of diodes are formed by the deposition of a metal on the surface of metal conductor.



3.**Light emitting diode (LED)**

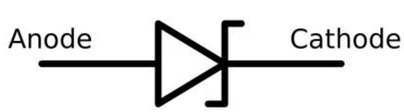
A light emitting diode (LED) is a semiconductor diode that emits light when electrical current is applied in the forward direction of the diode.

LEDs are made from variety semiconductor materials depending on the wavelength of the light required. The most used materials for the visible LEDs are gallium phosphide and gallium arsenic phosphide. Consider the figure below shows a light emitting diode and its symbol.



4**. Zener diode**

Zener diodes are specially manufactured diodes designed to be operated in the reverse breakdown voltage. Every zener diode is manufactured for a specific reverse breakdown voltage called the Zener diode. Consider the figure below shows a symbol of Zener diode.

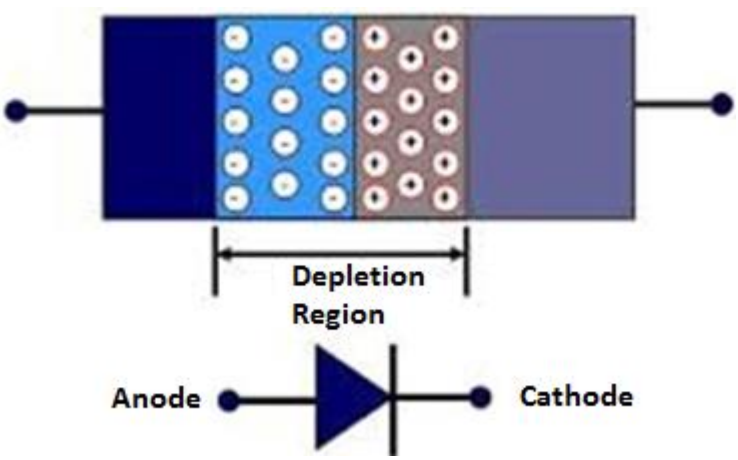


NB: Diode is designed to work in mode of forward and reversing biasing junction.

RECTIFICATION

This is the change of alternating current (alternate current to direct current the diode used to rectifier of A.C to D.C) to direct current has so many uses compare with an alternate current but is so expensive to produce a direct current also to reduce a cost is necessary to produce an alternate current then rectified to direct current.

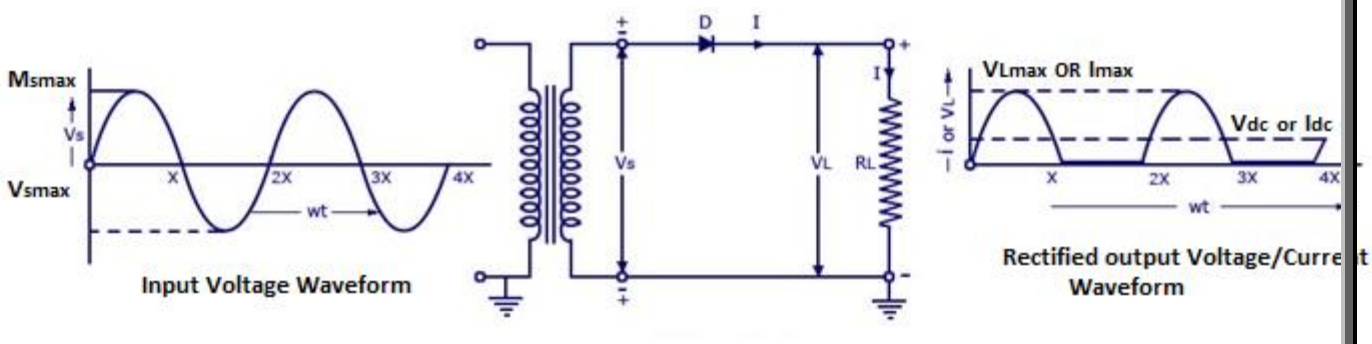
The conversion process is done through a p-n junction biased. A junction diode which is a biased p–n Junction is denoted by a circuit symbol.



The arrow head show the direction of flow of positive charges. It has been observed that p–n junction conducts current easily when is forward biased and practically no current flow when it is reversed biased. This outstanding property of the semiconductor for diode permit to be used as a rectifier i.e. it changes the alternating current to direct current.

1. **HALF WAVE RECTIFICATION**

In a half wave rectification only one diode is used. The diode conducts current only during the position half cycles of the input alternate current supply. During the negative half cycle of alternate current no current is conducted and hence no voltage appears across load.



Half-wave Rectifier

Half Wave Rectifier Operation

To understand the operation of a half wave rectifier perfectly, you must know the **theory part** really well. If you are new to the concepts of p-n junction and its characteristics, I recommend you to read the half wave rectifier theory part first.

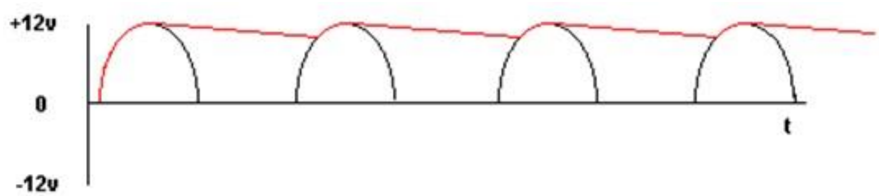
The operation of a half wave rectifier is pretty simple. From the theory part, you should know that a p-n junction diode conducts current only in 1 direction. In other words, a p-n junction diode conducts current only when it is forward biased. The same principle is made use of in a half wave rectifier to convert **AC** to **DC**.

The input we give here is an alternating current. This input voltage is stepped down using a transformer. The reduced voltage is fed to the diode **'D'** and load resistance **RL**. During the positive half cycles of the input wave, the diode **'D'** will be forward biased and during the negative half cycles of input wave, the diode **'D'** will be reverse biased.

We take the output across load resistor **RL**. Since the diode passes current only during one half cycle of the input wave, we get an output as shown in diagram. The output is positive and significant during the positive half cycles of input wave. At the same time output is zero or insignificant during negative half cycles of input wave. This is called **half wave rectification**. Before used a **d-c** current produced in half wave circuit must be smoothed.

The Smoothing Capacitor

We saw in the previous section that the single phase half-wave rectifier produces an output wave every half cycle and that it was not practical to use this type of circuit to produce a steady DC supply... We can therefore increase its average DC output level even higher by connecting a suitable smoothing capacitor across the output of the bridge circuit as shown below.



2. FULL WAVE RECTIFICATION

In full wave rectification current flow through the load in the same direction for both half cycles of the input alternate current voltage.

The following two circuits are commonly used for full wave rectification;

- Center – tap full rectification
- Full wave bridge rectification

1. CENTER TAP FULL RECTIFICATION

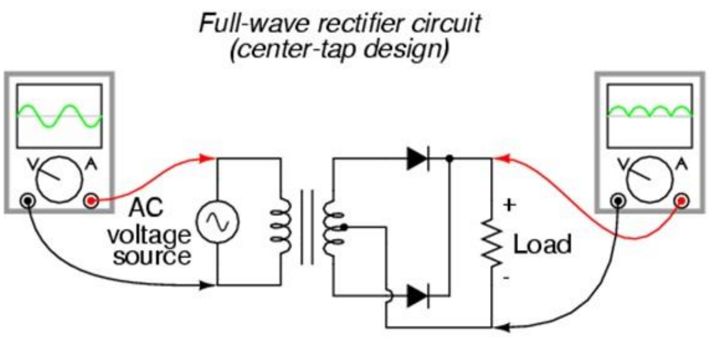
In this type the two diodes are used and the input is tapped at the center hence converted from alternate current to direct current form.

In a **central tap Rectifier** circuit two diodes are now used, one for each half of the cycle. A multiple winding transformer is used whose secondary winding is split equally into two halves with a common center tapped connection, **(C)**. This configuration results in each diode conducting in turn when its anode terminal is positive with respect to the transformer center point **C** producing an output during both half-cycles, twice that for the half wave rectifier so it is 100% efficient as shown below.

The full wave rectifier circuit consists of two power diodes connected to a single load resistance **(R_L)** with each diode taking it in turn to supply current to the load. When point **A** of the transformer is positive with respect to point **C,** diode **D_1** conducts in the forward direction as indicated by the arrows.

When point **B** is positive (in the negative half of the cycle) with respect to point **C**, diode **D_2** conducts in the forward direction and the current flowing through resistor **R** is in the same direction for both half-cycles. As the output voltage across the resistor R is the phase sum of the two waveforms combined, this type of full wave rectifier circuit is also known as a "bi-phase" circuit.

As the spaces between each half-wave developed by each diode is now being filled in by the other diode the average **DC** output voltage across the load resistor is now double that of the single half-wave rectifier circuit and is about **$0.637V_{max}$** of the peak voltage, assuming no losses.



The peak voltage of the output waveform is the same as before for the half-wave rectifier provided each half of the transformer windings have the same rms voltage value. To obtain a different **DC** voltage output different transformer ratios can be used. The main disadvantage of this type of full wave rectifier circuit is that a larger transformer for a given power output is required with two separate but identical secondary windings making this type of full wave rectifying circuit costly compared to the "Full Wave Bridge Rectifier" circuit equivalent.

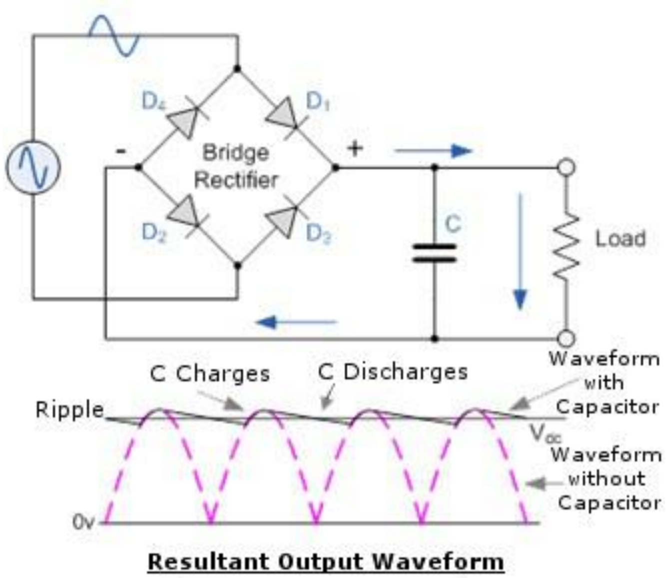
2**. FULL WAVE – BRIDGE RECTIFICATION**

The bridge – circuit produces full wave rectification without the use of center tapped secondary, it consists four diodes.

The Full Wave Bridge Rectifier

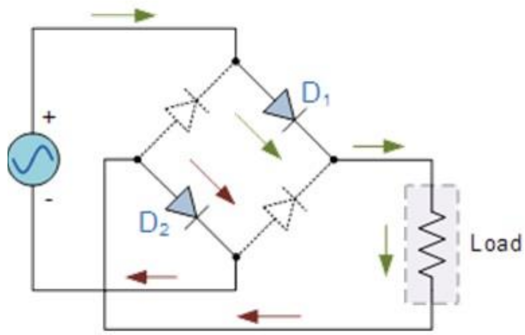
Another type of circuit that produces the same output waveform as the full wave rectifier circuit above is that of the **Full Wave Bridge Rectifier**. This type of single phase rectifier uses four individual rectifying diodes connected in a closed loop **"bridge"** configuration to produce the desired output. The main advantage of this bridge circuit is that it does not require a special center tapped transformer, thereby reducing its size and cost. The single secondary winding is connected to one side of the diode bridge network and the load to the other side as shown below.

The Diode Bridge Rectifier



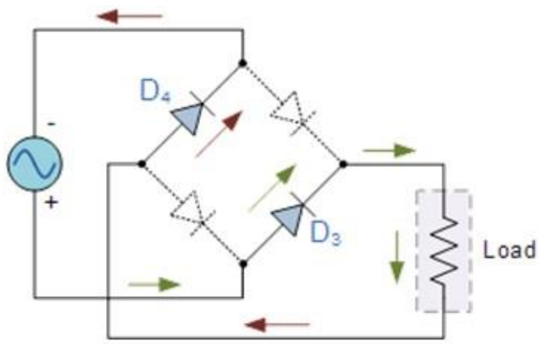
The four diodes labeled **D**$_1$ to **D**$_4$ are arranged in "**series pairs**" with only two diodes conducting current during each half cycle. During the positive half cycle of the supply, diodes **D1** and **D2** conduct in series while diodes **D3** and **D4** are reverse biased and the current flows through the load as shown below.

The Positive Half-cycle



During the negative half cycle of the supply, diodes **D3** and **D4** conduct in series, but diodes **D1** and **D2** switch "**OFF**" as they are now reversing biased. The current flowing through the load is the same direction as before.

The Negative Half-cycle



As the current flowing through the load is unidirectional, so the voltage developed across the load is also unidirectional, the same as for the previous two diode full-wave rectifier.

TRANSISTOR

A transistor is a semiconductor device used to amplify and switch electronic signals and electrical power. It is composed of semiconductor material with at least three terminals for connection to an external circuit. A voltage or current applied to one pair of the transistor's terminals changes the current through another pair of terminals. Because the controlled (output) power can be higher than the controlling (input) power, a transistor can amplify a signal. Today, some transistors are packaged individually, but many more are found embedded in integrated circuit.

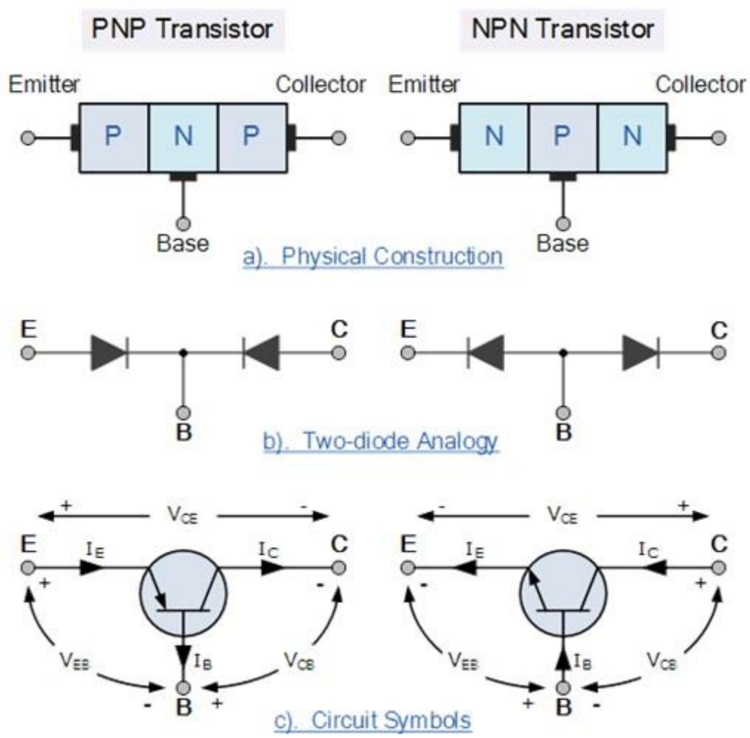
A bipolar junction transistor (BJT or bipolar transistor) is a type of transistor that relies on the contact of two types of semiconductor for its operation. BJTs can be used as amplifiers. Switches or in oscillators BJTs can be found either as individual discrete components, or in large numbers as parts of integrated circuits.

Bipolar transistors are so named because their operation involves both electrons and holes. These two kinds of charge carriers are characteristic of the two kinds of doped semiconductor material; electrons are majority charge carriers in n-type semiconductors, whereas holes are majority charge carriers in p-type semiconductors. In contrast, unipolar transistors such as the field effect transistor have only one kind of charge carrier.

Charge flow in a BJT is due to diffusion of charge carrier across a junction between two regions of different charge concentrations. The regions of a BJT are called emitter, collector, and base. A discrete transistor has three leads for connection to these regions. Typically, emitter is heavily doped compared to other two layers, whereas majority charge carrier concentrations in base and collector layers are about the same. By design, most of the BJT collector current is due to the flow of charges injected from a high-concentration emitter into the base where there are minority carriers that diffuse toward the collector, and so BJTs are classified as minority-carrier devices.

BJTs come in two types, or polarities, known as PNP and NPN based on the doping types of the three main terminal regions. An NPN transistor comprises two semiconductor junctions that share a thin p-doped anode region, and a PNP transistor comprises two semiconductor junctions that share a thin n-doped cathode region.

This is the dielectric device by joining either two N–type semiconductor sand mixed by one p–type OR two p–type send mixed by one N–type semiconductor. There are two types of transistor which are;



TERMINALS OF A TRANSISTOR

Base (B)- central point of a transistor.

Emitter (E)- line at which the electric signals flows into or out of the transistor. The line carrying signal in the input circuit.

Collector (C)- line at which the signal is taken from the transistor to the appliance i.e. the line taking signal from the transistor into

Emitter circuit is always in the forward BIAS mean while the collector circuit is in the reverse bias.

ELEMENTARY ASTRONOMY

Introduction

The word astronomy is derived from the Greek **astron**, meaning **"star"** and **nomos**, meaning "**laws or cultures**". The literal meaning of the word **astronomy** therefore is "**law of the stars**".

Astronomy is the branch of science that deals with study of the origin, evolution, composition, distance and the motion of all bodies (objects) and scattered matter in the universe.

The universe is the totality of space and time together with matter and energy.

People who are involved in astronomy are known as **astronomers**. Astronomers study the objects that compose the physical universe, stars, planets, moons, galaxies and nebulae.

Astronomy is one of the oldest fields of science. Early astronomy involved observing the regular patterns of the motions of visible celestial objects, especially, the sun, moon, stars and planets with their naked eyes. An example of astronomy was the study of the changing position of the sun along the horizon or the changing appearances of stars in the course of the year. These were used to establish agricultural and ritual calendars.

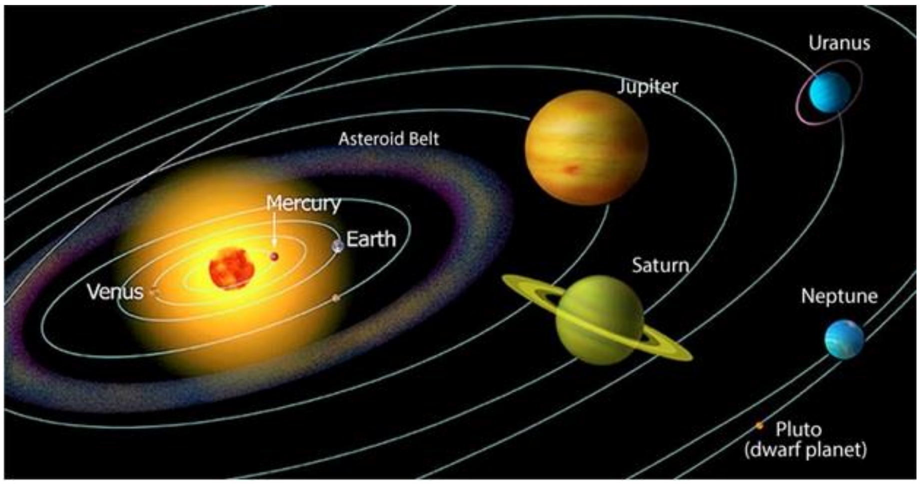
Importance of astronomy to mankind

Astronomy has been an important tool for thousands of years. The following are some ways in which astronomy is important;

1. It was the earliest method of measuring time. A day was the duration between sunrise and sunset while a month was derived from phases of the moon. The year was derived from the changing position of sunrise.
2. It was used to develop calendars that made it possible to predict the seasons. The season was very important in agriculture as they dictated the planting time and the harvesting time.
3. It was used in both land and sea navigation based on the knowledge of the position of the sun during the day and the stars at night.
4. Today, astronomy helps us to understand where the earth and the life it supports originated from and how it evolved.
5. Astronomy presents a new frontier for exploration.

The solar system

The solar system is made up of the sun arid the celestial objects bound to it by gravity. The objects include the **eight planets** and their known moons and billions of small bodies that include asteroids, comets, meteoroids and interplanetary dust.



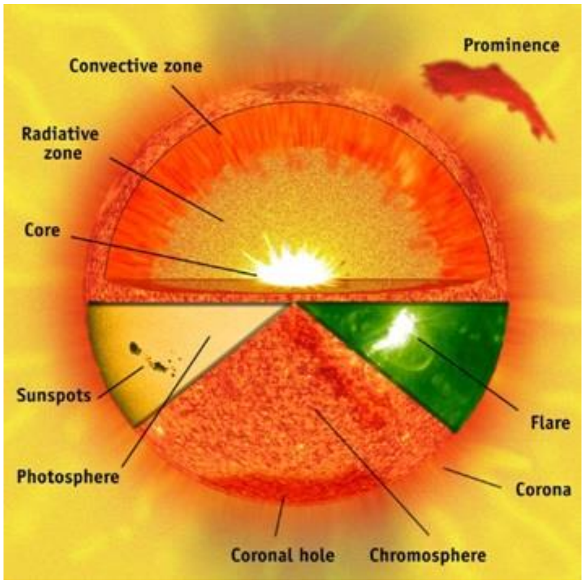
Solar system

Stars and planets

A star is a large celestial body made up of hot as known as plasma.

Plasma refers to an ionized gas in which a certain proportion of electrons are free rather than bound to an atom or molecule. Stars radiate energy derived from the thermonuclear reactions in the interior region. The sun is the largest star.

The sun is also the closest star to the earth at a mean distance of **149.60 million kilometers**. This distance is known as the **astronomical unit** and is used to measure distances across the solar system.



Structure of a star - the sun

A galaxy is a giant collection of stars, gas and dust. Most stars in the universe are in the galaxies. Nearly all of the stars visible in the night sky are within our own galaxy, our galaxy is called the Milky way galaxy.

A planet is a major (large) object which is in orbit around a star. There are eight such objects which are in orbit around the sun. These are commonly referred to as "the planets". They are Mercury, Venus, Earth, Mars, Jupiter, Saturn, Uranus and Neptune.

The following are the three defining characteristics of a planet:

1. It is a celestial body that orbits a star.
2. It is massive enough so that its own gravity causes it to assume a spherical shape.
3. It has cleared the neighborhood around its orbit.

NOTE: Because Pluto resides in an area of space populated by numerous other objects, it is no longer considered a planet. Pluto is now designated a dwarf planet. A dwarf planet does not meet the third characteristic, i.e. has not cleared the neighborhood around its orbit.

Differences between stars and planets

stars	Planets
Emit own light.	Do not emit their own light.
Twinkle at night.	Do not twinkle at night.
Appear to be moving from east to west.	Planets move around the sun from west to east.
Their temperatures are usually very high.	Their temperatures depend on their distances from the sun.
Countless in number.	There are eight in the solar system.
Very big in size but they appear small because they are very far away.	Very small in size as compared to stars.

Asteroids, sometimes called **minor planets**, are small solar system bodies in orbit around the sun, especially in the inner Solar System. Asteroids are smaller than planets but larger than a speck of dust.

A comet is a solid body orbiting the sun typically composed of rock dust, or ice. Most comets were formed from condensed interstellar gas and dust clouds in the early stages of the creation of the universe.

Gravitational force

The terms gravitational force and gravity are used often to explain the something, but there is a definite difference between the two terms.

Gravitational force is the attractive force existing between any two objects that have mass. Gravitational force pulls objects together.

Since gravitational force acts on all matter in the universe, from the largest stars to the smallest atoms, it is often called **universal gravitation**. Sir Isaac Newton was the first person to fully recognize that the force holding any object to the earth is the same as the force holding the moon, the planets, and other heavenly bodies in their orbits.

According to Newton's law of universal gravitation, every single point mass attracts every other point mass by a force directed along the line joining the two masses. The force is proportional to the product of the two masses and inversely proportional to the square of the distance between the point masses:

$$F=\frac{GM_1M_2}{r^2}$$

Where **F** is the magnitude of the attractive gravitational force between the two point masses; **G** is the universal gravitation constant; $\mathbf{m_1}$is the mass of the first point mass; $\mathbf{m_2}$ is the mass of the second point mass; and **r** is the distance between the centers of the two point masses.

Gravitational force is actually a very weak force. The pull is too weak to be felt between two people. It is only when one of the masses is very large to the size of a planet that the force of gravity can be felt.

The huge gravitational force of the nearest star, the sun, holds together the eight planets of the solar system. The planets move through space at speeds that just balance the suns gravitational pull, so they are locked into a permanent path (orbit) around the sun. Natural satellites (moons) orbit planets. Artificial (human made) satellites and spacecraft orbit the earth' in the same way as the moon orbits the planets.

Gravity denotes the gravitational force that occurs between the earth and other bodies. Gravity is the force acting to pull objects towards the earth. Gravity is the force that holds us on the ground and causes objects to fall back to the ground after being thrown up in the air.

The earth's gravitational pull extends out into space in all directions. The further you move away from, the center of the earth, the weaker the force becomes.

The measure of the force of gravity on an object on the earth's surface is the weight of that object Weight is measured in **Newton's (N)**. The weight of an object changes depending on its location in the universe.

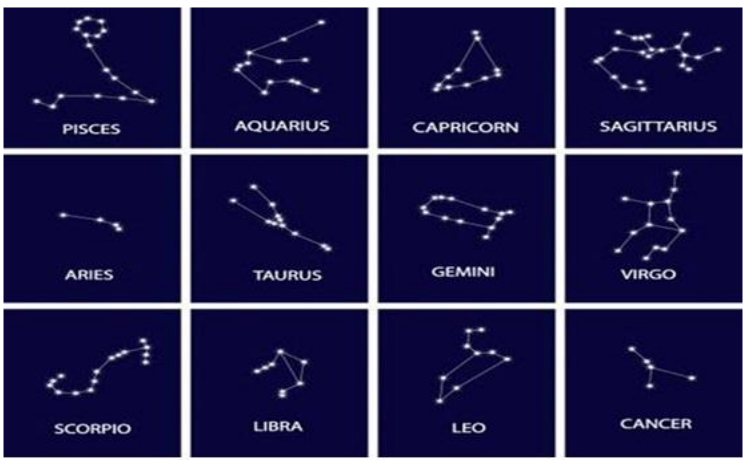
Constellations

A constellation: is a group of stars that form a definite shape or pattern when viewed from the earth.

Constellations are usually named after mythological characters, people, animals and things.

There are about **88** known constellations. The various constellations are visible during a particular period of the year.

Some examples of the known constellations are shown, in the table below;



Uses of constellations

The following are some uses of constellations:

1. **Religious**

In early days, people thought that the gods lived in the heavens and that the gods created the constellations. Many cultures believed that the positions of the stars were their god's way of telling stories. Indeed, the Greeks named the constellations after their mythological heroes and legends. For example, to the ancient Greeks, Orion was a great hunter. He was the son of Neptune (god of the sea).

2. **Agricultural**

Before there were proper calendars, people had no way of determining when to sow or harvest except by the stars. Constellations made the patterns of the stars easy to remember. The ancient peoples knew, for example, that when the constellation Orion started to be fully visible, winter was coming soon. The constellations allowed farmers to plan ahead.

3. **Navigation**

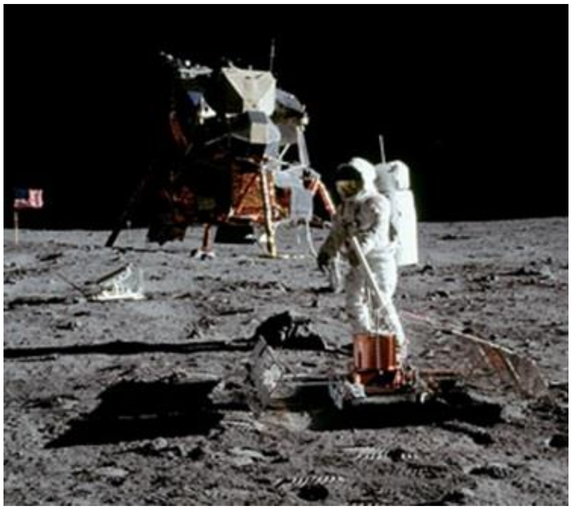
It is fairly easy to spot Polaris (The North Star) once you have found Ursa Minor (Little Dipper constellation). One can figure out his or her latitude (North or South) just by looking at how high Polaris appears in the night sky. This allowed for ships to travel across the globe. It allowed for the discovery of America, the spread of European culture and civilization as we know it today.

The earth and its moon form a unique pair in the solar system.

The moon of the earth is the **sixth largest** in the solar system. It has a diameter of **3476k**m and a mass of **7.35 x 10^{22} kg**.

Besides the earth, the moon is the only other body in the solar system upon which humankind has walked.

Astronaut Buzz Adrian on the moon with the Apollo 11 spacecraft.



Like the earth, the moon has an iron core surrounded by a rocky mantle and crust. Unlike the earth, no part of the moon's iron core is molten so it does not have a magnetic field. Surface gravity on the moon is about **1/6** that on the earth. An object weighing **120 kg** on the earth would only weight **20 kg** on the moon.

The moon revolves in an anticlockwise direction around the earth in an **elliptical orbit**. The moon's orbit is tilted at **5°** relative to the earth's orbit around the sun. The distance between the earth and the moon varies from perigee (nearest the earth) where it is **356,000km** to apogee (furthest from the earth), where it is **406,000km**. The average distance is 384 000 km. It takes the moon **27.3** earth days to complete one orbit, a period of time called the Sidereal month.

The moon also rotates about its axis at a rate equal to its rate of revolution. In other words while the moon is completing one orbit around the earth, it also spins round one time. The result is that the same side of the moon is always facing the earth. The side facing the earth is called **the near side** while the side that faces away is called **the far side.** The spinning of the earth causes the moon to rise and set each day, just like the sun. However, because of the moon's orbital motion around the earth, it (the moon) rises about 50 minutes later each day. As a result, the moon can be seen at different times of the day and night during a month.

Temperatures on the surface of the moon are on average **107°C** during the day and **-53°C** during the night.

Surface features of the moon

There are two primary types of terrain on the moon. These are the heavily cratered very old **lunar highlands** and the relatively smooth and Younger Maria.

From the earth, the moon's surface appears to have bright and dark regions when viewed with the unaided eye. The bright areas are the lunar highlands that have many craters and are covered with a highly reflective layer of fine dust. The highlands are geologically the oldest parts of the moon's surface.

The dark regions are low areas similar to ocean basins on the earth. They are filled with dark solidified lava and are less cratered than the highlands. Galileo called these **areas Maria**, the Italian word for seas, because their dark smooth surfaces appeared to be large bodies of water.

A photograph of the moon's surface showing highlands and Maria.

The Maria, which makes about **16%** of the moon's surface, are huge impact craters that were later flooded with molten lava. Most of the Maria {s covered with regolith, a mixture of fine dust and rocky debris produced by meteor impact.

Ocean tides

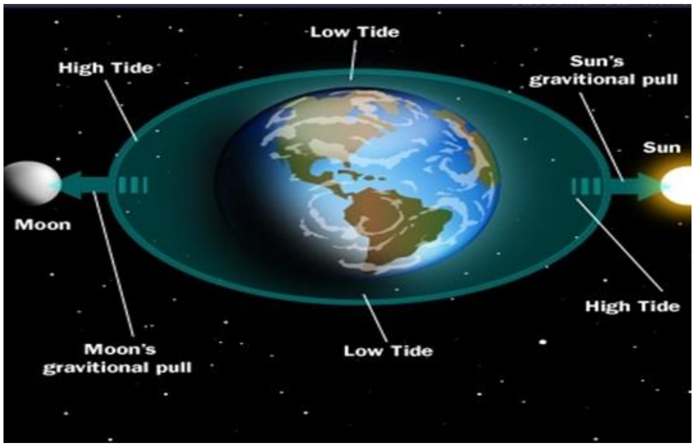
Tides are periodic rises and falls of large bodies of water. Tides are caused by the **gravitational interaction** between the earth and the moon.

The earth and the moon are attracted to each other, just like magnets are attracted to each other. The moon tries to pull at anything on the earth to bring it closer. But the earth is able to hold onto everything, except the water. Since the water is always moving, the earth cannot hold onto it and the moon is able to pull at it. This results in ocean tides. Each day, there are two high tide and two low tides.

The ocean is constantly moving from high tide to low tide, and then back to high tide. There is a time interval of about 12 hours and 25 minutes between the two high tides.

How tides occur

The gravitational attraction of the moon causes the oceans to bulge out in the direction of the moon. Another bulge occurs on the opposite side since the earth is also being pulled towards the moon (and away from the water on the far side). Ocean levels fluctuate daily as the sun, moon and earth interact. As the moon travels around the earth and as they together travel around the sun, the combined gravitational forces cause the world ocean water levels to rise and fall Since the earth is rotating while this is happening, two tides occur each day.



Types of tides

There are two main types of tides. These are **spring tides** and **neap tides.**

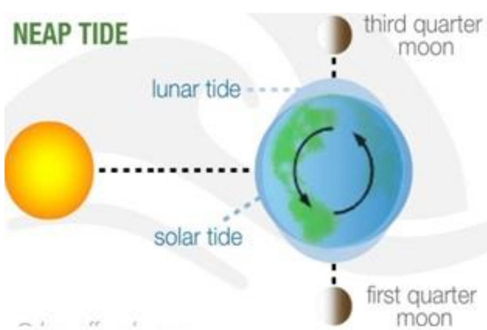
Spring tides

Spring tides occur during the full moon and the new moon. During this time, the earth, the sun and the moon are in a line. The gravitational forces of the moon and the sun both contribute to the tides. At these times, the high tides are very high and the low tides are very low. These are known as a spring high tide and a spring low tide, respectively. Spring tides are especially strong tides.

Proxigean spring tide is a. rare unusually high tide. The Proxigean spring .tide occurs when the moon is both unusually close to the earth (at its closest perigee, called the-proxigee) and in the new moon phase' (when the moon is between the sun and the earth). The proxigean spring tide occurs at most once every **1.5 year.**

Neap tides

When the sun and the moon are not aligned, the gravitational forces cancel each other out, and the tides are not.



Very high or very low, these are called neap tides.

Neap tides occur during quarter moons. During this time, the gravitational forces of the moon and the sun are perpendicular to one another (with respect to the earth). This causes the bulges to cancel each other. The result is a smaller difference between high and low tide and is known as a neap tide. Neap tides are especially weak tides.

Chapter summary

1. Astronomy is the study of the origin evolution, composition, distance and the motion of all objects and scattered matter in the universe.
2. In early times, astronomy was used to measure time, predict seasons and in navigation.
3. The solar system is made up of the eight planets, the sun, thousands of asteroids and countless comets and meteoroids.
4. Stars are giant spheres of hot gases called plasma.
5. A galaxy is a large group of stars, dust and gas held together by mutual gravitational forces.
6. Gravitational force is the attractive force that exists between any two objects that have mass.
7. Gravity is the gravitational force that exists between the earth and other bodies.
8. A constellation is a group of stars that form a definite pattern in the sky when viewed from the earth.
9. Constellations are used in religion, agriculture and in navigation.
10. The earth has one moon. The moon of the earth is the sixth largest in the solar system.
11. Interaction between the sun, the earth the moon and the ocean waters results in ocean tides.
12. There are two main types of ocean tides. These are spring tides and neap tides.

GEOPHYSICS

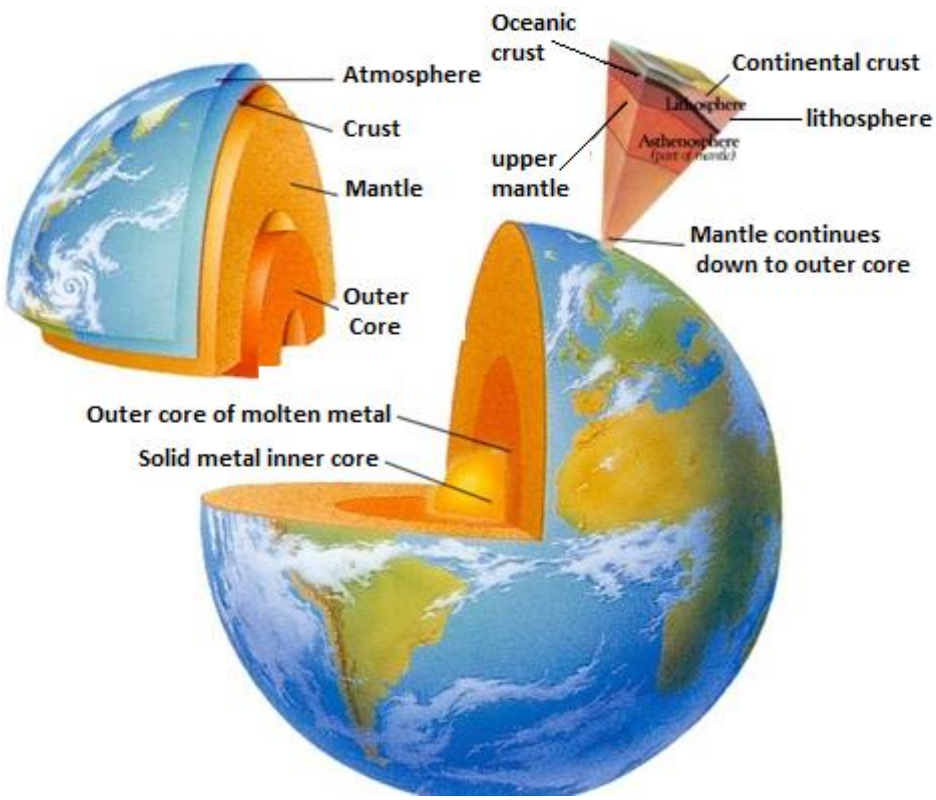
Introduction

Geophysics is the branch of science that is concerned with the physical, chemical, geological, astronomical -and other characteristic properties of the earth.

It deals with geological phenomena such as the temperature distribution of the earth's interior, the source, configuration and the geomagnetic field.

Interior structure of the earth

The structure of the earth is composed of three major zones arranged in a concentric manner. These are **crust**, **mantle** and **core**;



Internal structure of the earth

The crust

The crust is the outer solid layer of the earth. It is extremely thin (**5** to **15km**) compared to the radius of the earth (**6371 km**).

There are two types of crust, namely **continental crust** and **oceanic crust.**

Continental crust is heterogeneous and of relatively low density (**2 to 2.8 tonnes per cubic meter**). It is composed mainly of **granites** and **sedimentary rock**.

Oceanic crust is basaltic and has a. higher density (**3.0 to 3.1 tonnes per cubic meter**). Both the continental and oceanic crusts float on the denser mantle. Because of its low density, the continental crust floats on the mantle at a higher elevation, forming the land masses and mountains. The continental crust is **30 to 70km** thick.

The dense oceanic crust floats at a lower elevation forming oceanic basins. It is about **8km** thick.

The boundary between the crust and the mantle is called **Mohorovicic discontinuity** or **simply Moho**. It is a zone between one and several kilometers thick.

The mantle

The mantle begins from the Moho and extends to a depth of **2900km** below the earth's surface, up to its boundary with the earth's core. This boundary is called the **Gutenberg discontinuity**.

The mantle contains about **70%** of the earth's mass. It is made up of rocks, both in solid and in molten states. These rocks are said to be in a plastic state. The upper part of the mantle has a temperature of about **870°C**. The temperature increases downwards through the mantle to about **2200°C** near the core.

Circulation of materials in the mantle is the main mechanism of heat transfer from the core of the earth to the outer regions of the earth. It is the main force that drives the movement of continents as well as volcanism and earthquakes.

The core

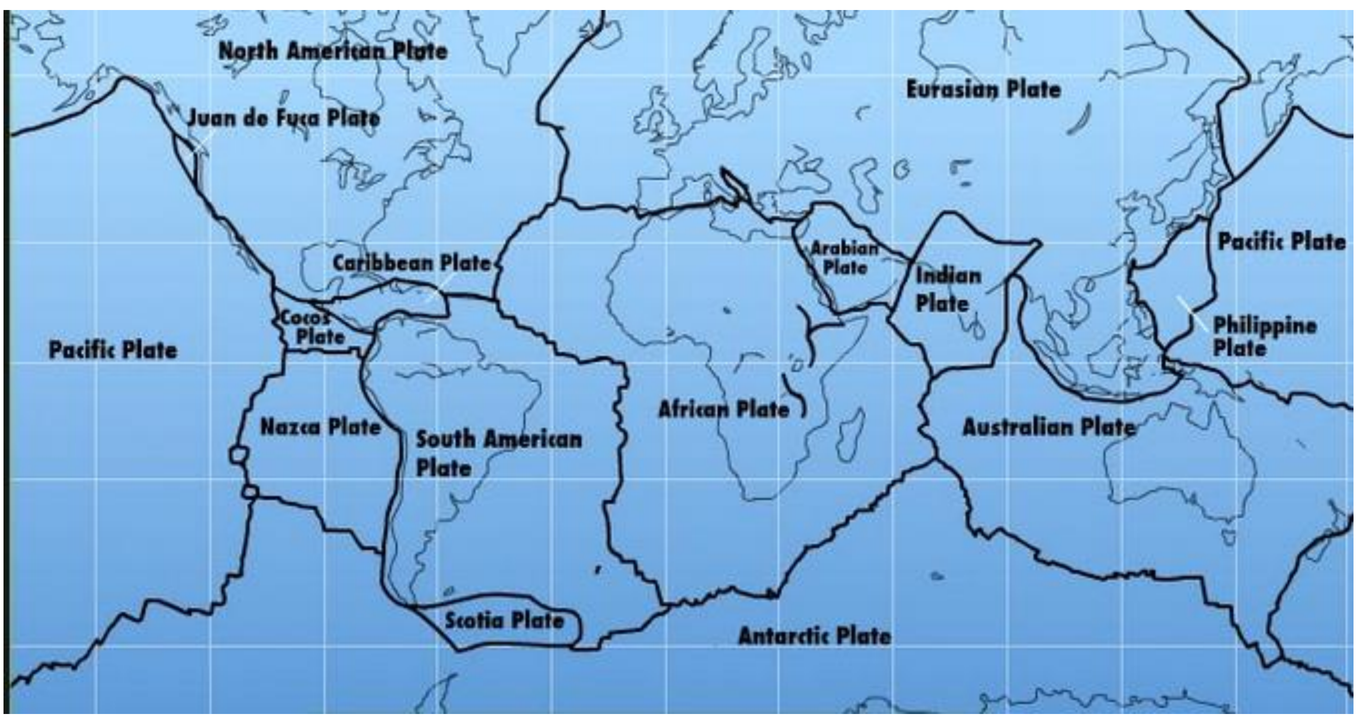
The core is the innermost part of the earth. It extends from the Gutenberg discontinuity to the geometric center. The core consists of two distinct regions, namely the **outer core** and the **inner core**.

The outer core is composed of a liquid of molten nickel and iron known as magma. It extends from the mantle to a depth of about **5000km** below the earth's surface. The inner core is solid; it is composed of iron-nickel alloys.

The material of the inner core is solid because of the high pressure at this depth.

Tectonic plates

The earth's crust and part of the mantle are cracked into huge pieces called **tectonic plates**. These plates float on top of the semi-molten rock underneath. They move about at a very slow speed. The movements of the tectonic plates mean that some continents are moving apart and some are moving towards each other. This process is referred to as **continental drift** and has been going on for hundreds of millions of years. Tectonic plate movements have split the continents as- we, know them today.



Tectonic plate

The line where two tectonic plates meet is called a **boundary**. There are three main types of boundaries. These are **destructive boundarie**s, **constructive boundaries** and **conservative boundaries**.

Destructive boundary (convergent)- is one found at the edges of two plates moving towards each other.

Constructive (boundaries divergent)- are formed at the edges of two plates moving away from each other.

Conservative boundaries- are formed when two plates slide past each other without moving apart or towards each other.

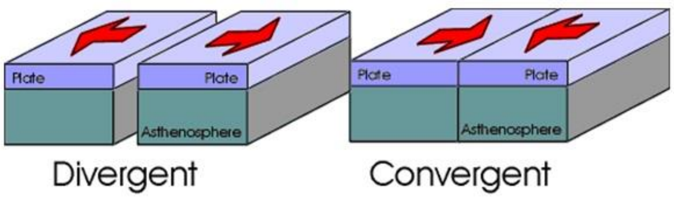


Diagram constructive and destructive boundary

Volcanoes and earthquake

Volcanoes and earthquakes are closely related. They are both caused by the movement of molten rock and heat deep inside the earth. These movements are referred to as **subterranean movements**. Most earthquakes and volcanic activity happen near tectonic boundaries.

Volcanoes

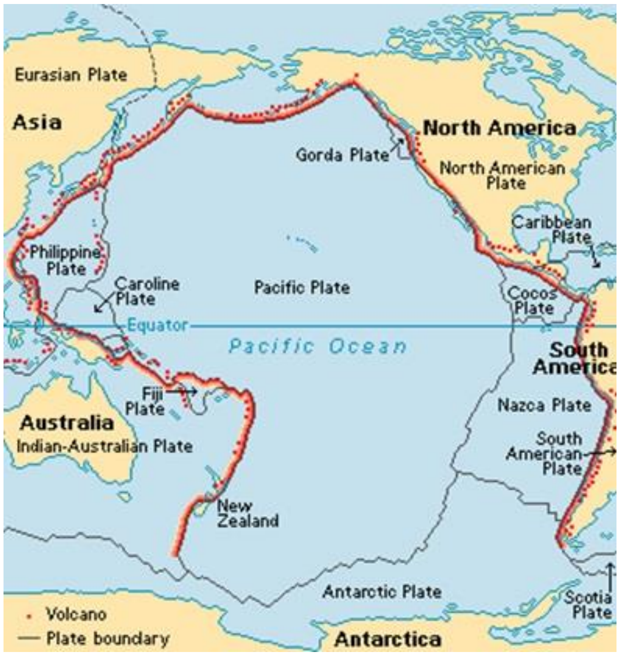
Volcanoes are places where molten rock called **magma** leaks out through a hole or a crack in the earth's crust. Magma originates from the mantle where high temperatures and pressure cause the rock to melt. When a large pool of magma is formed, it rises through the denser rock layer towards the earth's surface. Magma that has readied the earth's surface is called **lava**.

Volcanic eruption

Most volcanoes form along constructive and destructive boundaries between tectonic plates. However, few form far from plate boundaries**.**

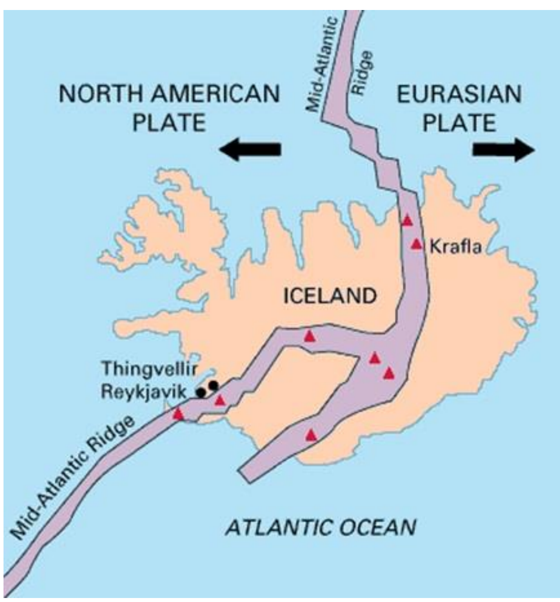
Volcanoes at destructive boundaries

When an ocean plate plunges under another plate, the ocean plate rubs against the plate above it and gets hot. The rock melts resulting in magma under the upper plate. This pool of magma forces its way through weak points in the crust. This creates a line of volcanoes parallel to the boundary but off to one side in the upper plate. Most of the world's volcanoes occur at destructive boundaries. Nearly all the way around the Pacific Ocean is a line of destructive plate boundaries whew ocean plates slide under continental plates and other ocean plates. These boundaries hare created a circle of volcanoes around the rim of the Pacific. This circle is called the **Ring of Fire**. It runs along the west coasts of South America, and North America, through Japan, the Philippines and New Zealand.



The ring of fire

Volcanoes can also form along constructive boundaries. These are formed where two plates move apart at the boundary. Magma moves up from underneath to fill the gap left by the separating plates. The Mid-Atlantic Ridge that runs along the floor of the Atlantic Ocean is a continuous volcano that is thousands of kilometers long. It formed at a constructive boundary between the plates.



The mid Atlantic ridge

Hot-spot volcanoes

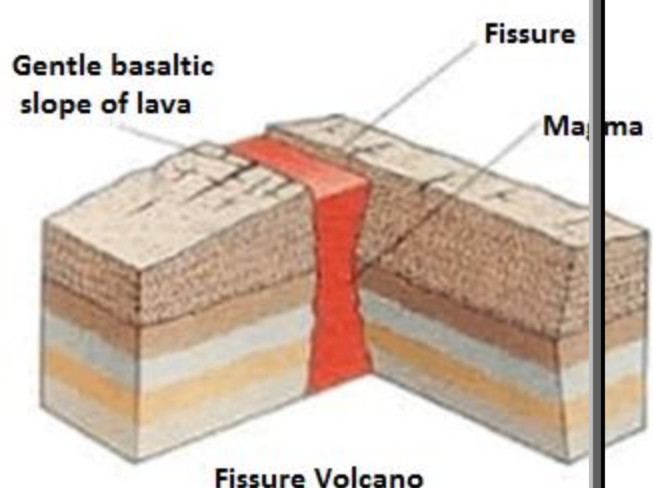
Volcanoes also erupt thousands of kilometers away from tectonic plate boundaries. It is thought that these eruptions occur over places in the mantle that are hotter than normal. Magma from these hot-spots forces its way through the crust above and onto the earth's surface. Nyamulagira Volcano in Congo is a hot-spot volcano.

Types of volcanoes

There are two main types of volcanoes, namely **fissure** and **central volcanoes**;

Fissure volcanoes

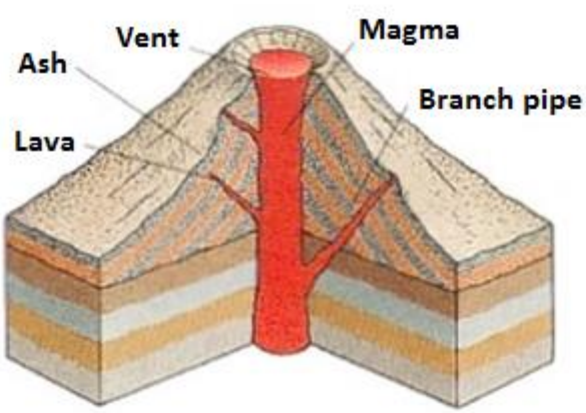
These occur along cracks in and between tectonic plates. They can be many kilometers long. Lava is usually ejected quietly and continuously, forming enormous plains or plateaus of basaltic volcanic rock.



Fissure volcano

Central volcanoes

These hare a single vertical main vent through which magma reaches the earth's surface. They usually develop a cone shape that builds up from successive layers of lava and ash.



Central volcano

Classification of volcanoes

Volcanoes are classified into three categories based on their frequency of eruption. These are **active**, **dormant** and **extinct volcanoes**.

Active volcanoes-Are those that either erupt constantly or have erupted in recent times. They include the Ol Donyo Lengai Volcano.

Dormant volcanoes-Are those that have been inactive for some time (a few thousand years) but can erupt again. An example is Mt. Kilimanjaro.

Extinct volcanoes-Are those that have not erupted in recorded history. They will probably never erupt again.

Effects of volcanoes

1.**Landscape** - Volcanoes have a great effect on the landscape. Much of the earth's surface is covered with volcanic rocks. Volcanoes are also responsible for the formation of many mountains and islands.

2.**Vegetation and wildlife** - Volcanic eruptions destroy vegetation. The eruptions sometimes set the surrounding vegetation on fire. Such fires consume huge tracts of vegetation that include forests, woodlands and grasslands.

3.**Wild animals-** are also killed by being buried in the lava or being burnt by the forest fires.

4.**Environment** - Besides the destruction of vegetation, volcanic eruptions emit harmful gases into the environment. Such gases include sulphur dioxide. Some of the gases contribute to global warming and climate change.

5.**Human life and property** - Volcanic eruptions sometimes kill people and destroy property. People who monitor volcanic activities usually warn people of an impending eruption so that people life can vacate such areas. However, some eruptions happen unexpectedly. Such eruptions bury people, animals and buildings in mountains of lava.

6.**Soil**- Volcanoes help in soil formation by bringing important soil minerals from deep underground onto the earth's surface.

7.**Minerals=** Volcanic eruptions also bring valuable minerals to the earth's surface. The minerals are important economic resources.

Earthquakes

Earthquakes happen when rocks in the earth's crust move suddenly, shaking the earth. About **10,000** earthquakes happen, every year but most are so small that they can only be detected by very sensitive instruments.

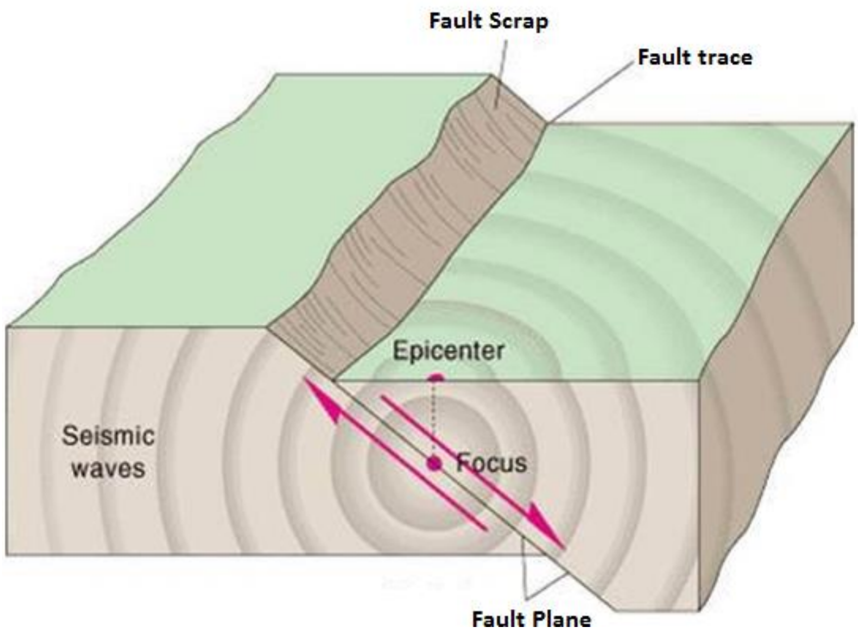
Earthquakes also occur as a result of movement of magma at constructive boundaries, under volcanoes and where continental plates collide and push mountain ranges.

How earthquakes occur

Earthquakes mostly occur on or near the boundaries between tectonic plates. However, earthquakes can also occur far from plate boundaries. Such earthquakes probably occur as a result of faults formed millions of years ago.

Actually, most earthquakes occur on or near destructive and conservative boundaries of tectonic plates.

Tectonic plates grind past each other, rather than slide past each other smoothly. As the plates move past each other they can become locked together due to friction. For some tune, they don't move and energy builds up.



Earthquake build-up

Pressure builds between them until the frictional force holding the plates together gives way. The plates move suddenly, releasing the pressure (energy) and then hold together again. This sudden jerk is what is felt as an earthquake.

An earthquake is a sudden motion or shaking of the earth caused by a sudden release of energy that has accumulated within or along the edges of the earth's tectonic plates.

The point within the earth where an earthquake begins is called the **hypocentre** or the focus of the earthquake.

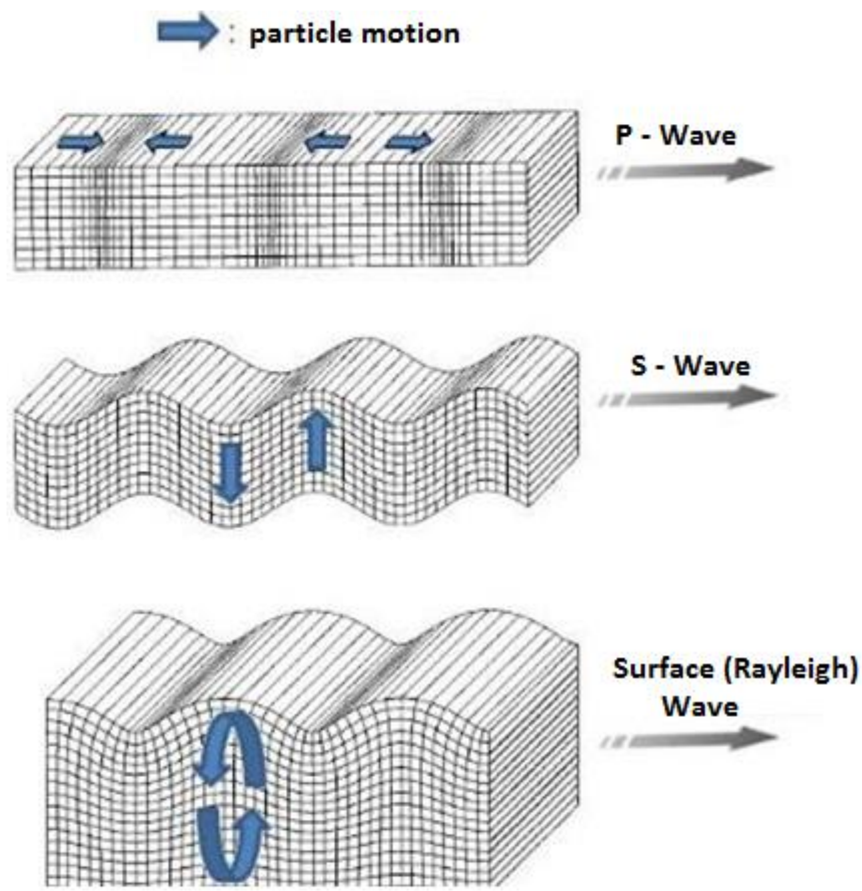
Earthquakes rarely occur along constructive plate boundaries.

Seismic waves

Earthquakes release their energy in three waves of energy called **seismic waves**. These are **primary waves, secondary waves** and **surface waves.**

Primary waves or p-waves are the-first waves released from the hypocenter. Primary waves are felt as a sudden jolt.

Secondary waves or s-waves arrive a few seconds later. They are felt as a series of side-to-side tremors.



Seismic waves

Surface wares radiate outward from a point on the earth's surface directly above the hypocenter. This point is called the **epicenter** of the earthquake. There are two types of surface waves. These are **Rayleigh waves** and **love waves**. Rayleigh waves create a rolling movement that makes the land surface move up and down.

Love waves make the ground shift from side to side.

It is the surface waves that do damage to surface structure such as buildings and hydroelectric power plants.

EARTHQUAKE SCALES

The nature of an earthquake is usually described by measuring two properties. These are the magnitude and intensity.

The magnitude of an earthquake is a. measure of the energy it releases. It is usually measured on the Richter scale. The Richter scale is based on the amplitude of the largest seismic wave recorded for an. earthquake, no matter what type of wave was the strongest.

The Richter scale magnitudes are based on a logarithmic scale (base 30). This means that for every whole number you go up on the Richter scale, the amplitude of the ground motion goes up ten times. Using the Richter scale, art earthquake .of magnitude **7.0** would result in ten' times the level of ground shaking as an earthquake of magnitude **6·0**.

The Richter scale can be used to describe earthquakes so small that they are expressed in negative numbers. The scale has no upper limit.

The intensity of an earthquake is a measure of its strength based on the changes it causes to the landscape. The intensity of an earthquake is usually measured on the Modified Mercalli scale. The scale is calibrated from 1 to 12. On this scale, level 1 is a minor tremor that causes no damage whereas level 12 causes total devastation. The table below shows a description of the **12 levels** on the Modified Mercalli scale.

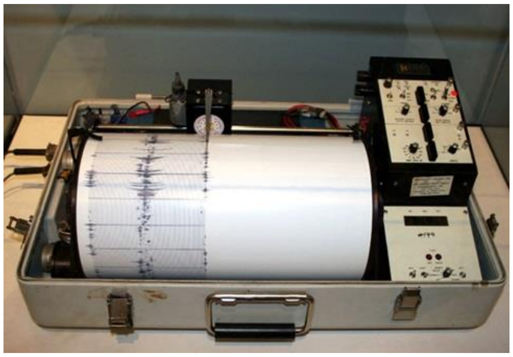
Intensity level	**Effect**
I	No felt, except by very few people under especially favorable condition.
ii	Felt only by o few people at rest, especially on upper floors of buildings. Delicately suspended objects may swing.
Iii	Felt quite noticeably by persons indoors, especially on upper

	floors of buildings. Many people do not recognize it as an earthquake. Standing motor cars may rock slightly. Vibration similar to that of a passing truck. Duration can be estimated.
iv	Felt indoors by many, outdoors by few during the day. At night, some awakened .Dishes, windows, .doors disturbed. Walls make cracking sound. Sensation similar to that of a heavy truck striking a building. Stationary cars rocked noticeably.
V	Felt nearly by everyone. Many are awakened. Some dishes, windows broken. Unstable objects overturned. Pendulum clocks may stop.
Vi	Felt by everyone. Many are frightened. Some heavy furniture moved. A few instances of fallen plaster. Damage slight.
Vii	Damage negligible in buildings of good design and construction. Slight to moderate damage in well-built ordinary structures. Considerable damage in poorly built or badly designed structures. Some chimneys are broken.
Viii	Damage slight in specially designed structures. Considerable damage in ordinary substantial buildings with partial collapse. Damage great in poorly built structures. Fall of chimneys, factory stacks, columns, monuments and walls. Heavy furniture overturned.
ix	Damage considerable in specially-designed structures. Well-designed frame structures thrown out of plumb. Damage great in substantial buildings, with partial collapse. Buildings shifted off foundations.
x	Some well-built wooden structures destroyed. Most masonry and frame structures destroyed with foundations. Rail bent.
xi	Few, if any structures (masonry) remain standing. Bridges destroyed. Rails bent greatly.
xii	Total damage. Lines of sight and level are distorted. Objects thrown into the air.

Note that: An earthquake can only have one magnitude. However, its intensity reduces as the seismic waves spread out from the hypocenter, just the same way the loudness of a sound changes as you move away from the source.

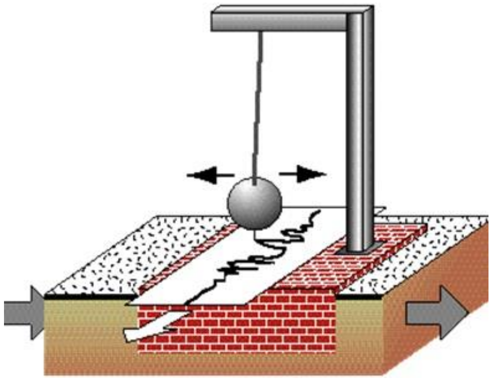
The seismograph

The seismograph is an instrument used to **record ground movements** caused by earthquakes. It measures ground oscillations-by recording the relative motion between a pendulum and the ground. It is also possible to use the ratio between the deflection of the pendulum and the acceleration of the ground to record an earthquake. The time of Initiation of ground oscillations is recorded and marked and are included on the graphs every minute and hour on the seismograph paper.



Seismograph

In order to measure ground motions, the seismograph must remain steady when the ground moves. Various types of pendulums have been used to obtain this steady state. The simplest type of pendulum is a heavy mass suspended by a wire or rod from a fixed point. , Other forms are the inverted and horizontal pendulum. The inverted pendulum has a heavy mass fixed to the upper end of a vertical rod pointed at its lower end, while the horizontal pendulum has a rod with a mass on its end which is suspended at two points so that it swings on a horizontal plane.



Vertical a pendulum seismograph

Recording the pendulum motion

The recording of the motion of the pendulum can be done, in various ways. The most common ones are the mechanical method, the optical method and the electronic method.

Mechanical method

In the mechanical method, a sheet of smoked paper is wrapped around a rotating drum and mounted to move with the earth. A moving pen connected to the pendulum presses lightly on the paper. As time passes, the drum rotates so that the recorded lines are not superimposed on each other. Deflection of the pendulum is commonly magnified mechanically by single or double multiplying levers so that the graph is easier to see.

This method is simple and economical. However, the seismograph must have a heavy mass to overcome the friction between the pen and the paper. Consequently, some mechanical seismographs weigh one tonne or more.

Optical method

The optical method still uses a pendulum motion to record the ground movements'. However, to overcome friction, mirrors are used to reflect the light onto photosensitive paper wrapped on a drum.

Electronic method

Technological developments have given rise to high-precision seismometers and sensors of ground motion. In these electromagnetic instruments, a coil is fixed to the mass of a pendulum and moves in a magnetic field. The electric current, generated in the coil, operates a galvanometer In the .same way a dynamo operates a motor. The voltages .produced by motions of the pendulum is passed through electronic circuits to amplify the ground motion for more exact readings.

The seismograph records both the magnitude and the intensity of the earthquake.

Earthquake hazards

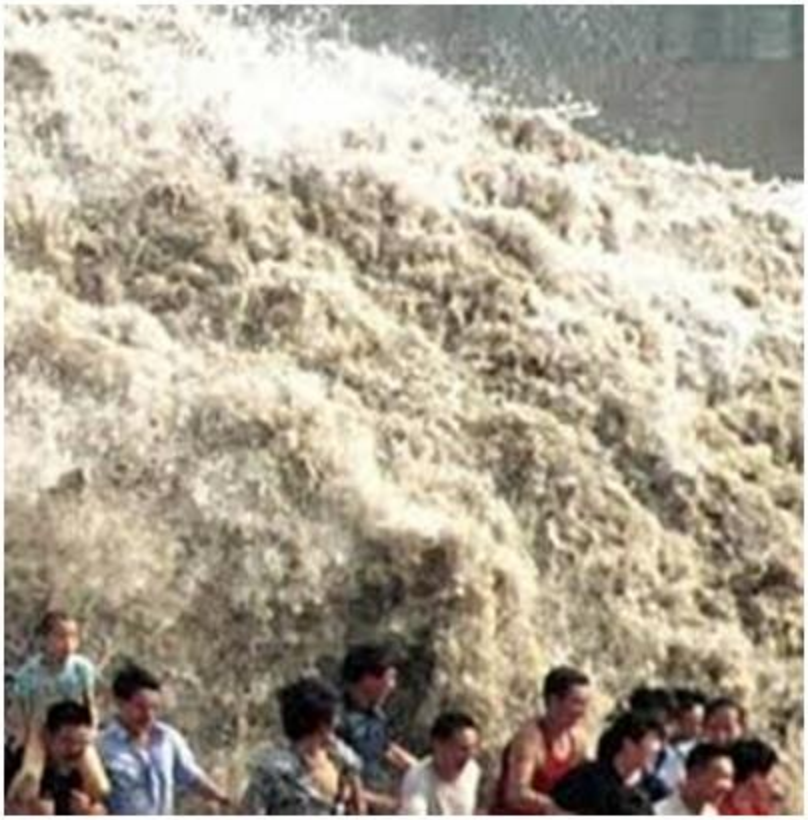
Earthquakes give rise to a number of hazards which pose a great risk to human life, animals, property and the environment at large. The following are some of the hazards associated with earthquakes:

(a) **Landslides** - The shaking caused by earthquakes can cause unstable hillsides, mountain slopes and cliffs to move downwards, creating landslides. In massive landslides created by earthquakes, soil and rock accelerate down the slope, sweeping away everything in their path. Landslides can fill valleys, creating temporary dams. These dams release floodwater when they collapse.

Earthquakes can also trigger avalanches on snow slopes

Land slide caused by an earth quake

(b)**Tsunamis** –If an earthquake happens in rocks under the sea or ocean, the shock waves disturb the water. The ocean floor can also rise or fall causing the water to rise or fall too. These movements create huge water waves called tsunamis that travel across the ocean. When a tsunami reaches shallow water, it slows down, its wavelength reduces and its height grows. When the tsunami hits the shore it crashes inland carrying everything in its way, including destroyed building.



(c)**Collapsing buildings** - Earthquakes do not actually kill people. It is the hazards that are associated with earthquakes that kill people. The majority of people killed or injured in earthquakes are trapped in buildings that collapse because of the ground, underneath shaking.

A strong earthquake can flatten a whole city. An example is the Japanese city of Kobe which was completely flattened by an earthquake measuring **7.2** on-the Richter scale. The earthquake occurred in **January 1995**. More than 6 000 people died and about **200,000** buildings collapsed or were damaged.

Collapse building

(d) **Fire outbreak** - An earthquake can trigger a fire outbreak. This happens when the earthquake causes gas or oil pipes to break. It can also occur as a result of the collapse of electricity lines.

(e) **Backward rivers** - Tilting ground can also make rivers change their course. This can result in the creation of earthquake lakes that cover huge tracts of previously settled land.

Earthquake warning signs

The following are important signs that are observed before an earthquake occurs:

1. Thermal indicator - A few months before the occurrence of an earthquake, the average temperature of the area keeps increasing. On the day of the earthquake, the temperature of a place is about **5** to **9 degrees Celsius** above the average normal temperature for that day.

2. Water indicator - About one or three days before an earthquake, there is- a sudden rise or fall in water levels in wells. The rise can be as high as one meter. The well water may turn muddy/. At times a fountain appears inside the well. Sometimes a fountain may appear in the ground. This normally happens; a few hours before the quake. There is also a sudden and rapid increase or decrease of water flow in the rivers. This happens about one to two days before the quake.

3. Seismo electromagnetic indicator -Before the occurrence of an earthquake the sub-surface temperature rises. As a result of this, the geomagnetic field is reduced. The reduction in geomagnetic field adversely affects the propagation of electromagnetic waves. This is experienced abundantly on the radio, television and telephone. This is a very reliable indicator. It is usually recorded about 10 to 20 hours before the quake.

If a particular radio station is received at a frequency of 1000kHz, the same station will be received in the potential epicenter area at higher frequencies, about 10 to 30 hours before the earthquake. Similarly, reception of television signals is affected.

The mobile telephone is one of the most reliable indicators of an impending earthquake. About 100 to 150 minutes before the occurrence of an earthquake, mobile telephones stop functioning or become malfunction.

Note that all the above indicators are valid only when seen and manifested extensively. Failure of one or two instruments should not be taken as an earthquake indicator.

4. Animal indicator - Between 10 and 20 hours before the occurrence of an earthquake, the entire animal kingdom becomes highly disturbed and restless. They move in a directionless manner and in fear. Birds do not perch on trees but move about at a low height, emitting a shrill noise. Rodents like rats and mongooses go into a panic. Domestic animals like cows, dogs and cats struggle against being tied up and may even turn on the owner.

5. Human indicator - Doctors and nurses have observed that some sensitive patients in hospitals become highly disturbed before an earthquake. They exhibit a sudden rise in blood pressure, heart trouble, headache, migraine and respiratory disorders. Indeed, the number of outpatients in< hospitals increases by five to seven times, some 10 to 20 hours before the quake.

The best human earthquake indicator' is the number of child deliveries in any hospital. On the penultimate day of the earthquake, the number of deliveries goes up about three to five times, while on the day of the earthquake it is as high as seven to eight times the average.

Precaution to be taken during an earthquake

The following are some precautions that can be taken to minimize injuries to or death of human beings, in the event of an earthquake:

1. If you are indoors during an earthquake, drop, cover and hold on. Get under a desk, table or bench. Hold on to one of the legs and cover your eyes. If there is no table or desk nearby, sit down against an interior wall. An interior wall is less likely to collapse than a wall on the outside shell of the building.

2. Pick a safe place where things will not fall on you away from windows or tall, heavy furniture.

3. Do not run outside when an earthquake happens because bricks, roofing and other materials may fall from buildings during and immediately after an earthquake, injuring persons near the building.

4. Wait in your safe place until the shaking stops, then check to see if you are hurt. You will be better able to help others if you take care of yourself first, and then check on the people around you.

5. Move carefully and watch out for things that have fallen or broken. Creating hazards. Be ready for additional earthquakes called aftershocks.

6. Be on the lookout for fires. Fire is the most common earthquake-related hazard due to damaged gas and electrical lines.If you must leave a building after the shaking stops, use the stairs and not the elevator.Earthquakes can cause fire alarms and fire sprinklers to go off. You will not be certain whether there is a real threat of fire. As a precaution, use the stairs.

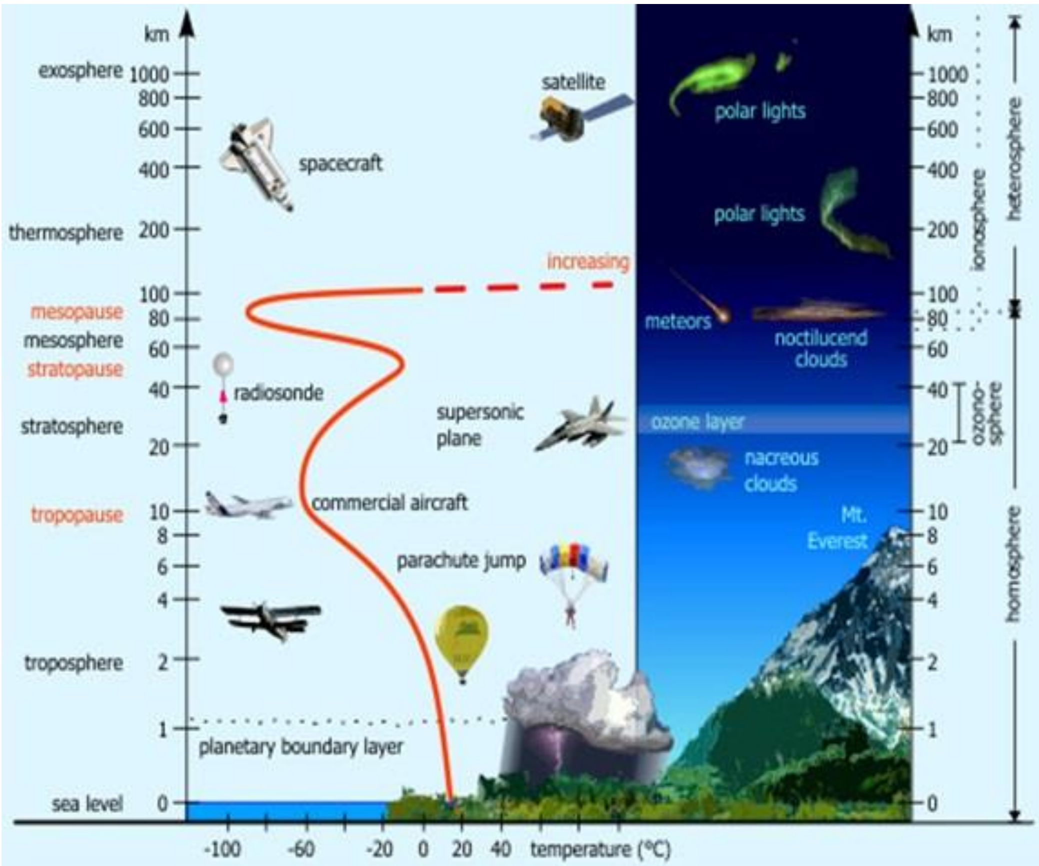
7. If you are outside during an earthquake stay outside. Move away from buildings, trees, streetlights and power lines. Crouch down and cover .your head. Bricks, roofing and other materials can fall from buildings, injuring persons nearby. Trees, streetlights and power lines may also fall, causing damage or injury.

Structure and composition of the atmosphere

The earth is surrounded above it by a layer of gases containing numerous small suspended solid and liquid particles. This layer is called the atmosphere. The atmosphere consists largely of, a mixture of gases, extending to a height of many kilometers above the earth. It has no outer boundary. It just fades into space. The dense part of atmosphere lies within 30 km above the earth's surface.

Structure of the atmosphere

The atmosphere is divided into regions based on its thermal characteristics (temperature changes), chemical composition, movement and density. The five regions are the troposphere, stratosphere, mesosphere, thermosphere and exosphere.



Vertical structure of atmosphere

Troposphere

This is the region nearest to the earth. It extends to an altitude of up to 10 km above the poles and 20 km above the equator.

This region is the densest part of the atmosphere. It contains 80% by mass of the atmosphere. It contains most of the atmosphere's water vapor. The temperature in this region decreases with altitude at an average rate of 6°C/km.

The troposphere is well mixed. Air molecules can travel to the top of the troposphere and back down again in just a few days. This mixing encourages changing weather.

Most weather phenomena occur in the troposphere. Clouds and rain are formed within this region.

The boundary which separates the troposphere and the stratosphere is called the tropopause. At the tropopause, the temperatures stop decreasing with height and become constant. The tropopause has an average height of 10 km.

Stratosphere

The stratosphere starts from the tropopause and extends to 50 km high. This layer is more stable, drier and less dense compared to the troposphere. The temperature in the stratosphere slowly increases with altitude.

The temperature increase is due to the presence of the ozone layer which absorbs ultraviolet rays from the sun. The ozone layer lies in the middle of the stratosphere between 20 and 30 km. Ozone is a triatomic (three-molecule) form of oxygen.

This layer plays the important role of absorbing ultraviolet radiations which would otherwise reach the earth's surface. Ultraviolet radiation is harmful to both animal and plant life on earth. The stable air of the stratosphere also prevents large storms from extending much beyond the tropopause.

Planes also fly in the stratosphere. This is because it has strong steady horizontal winds which are above the stormy weather of the troposphere.

The troposphere and stratosphere are collectively known as the lower atmosphere.

The boundary which separates the stratosphere and the other layers is called stratopause.

Mesosphere

The mesosphere starts just above the stratosphere and extends to 85 km high. The temperature at this layer decreases with altitude. The lowest temperature of the atmosphere (-90°C) occurs within this region.

The mesosphere is the layer in which most meteors burn while entering the earth's atmosphere.

The boundary which separates the mesosphere and the thermosphere is called the mesopause.

Thermosphere

This layer starts just above the mesosphere and extends up to 690 km high. The temperature increases with increasing altitude due to the sun's heat. The temperature in this region can go as high as 1 727°C. Chemical reactions occur much faster here than on the surface of the earth. This layer is also known as the upper atmosphere.

The lower part of the thermosphere, from 80 to 550 km above the earth's surface, contains the ionosphere. This is a region containing a high concentration of charged particles called ions and free electrons. The large number of free electrons in the ionosphere allows the propagation of electromagnetic waves.

The ionosphere also absorbs dangerous radiation. The radiation absorbed in the ionosphere includes hard and soft X-rays and extreme ultraviolet (EUV) radiation.

The ionosphere plays an important role in communications. Radio waves can be reflected off the ionosphere allowing radio communications over long distances.

Exosphere

The exosphere is the outermost region of the atmosphere. In this region, the atmospheric gas pressure is very low. Light atoms such as hydrogen and helium may acquire sufficient energy to escape the earth's gravitational pull

The upper part of the exosphere is called magnetosphere. The motion of ions in this region is strongly constrained by the presence of the earth's magnetic field.

This is the region where satellites orbit the earth.

Importance of the atmosphere

The following are some ways in which the layers of the atmosphere are important:

1. The troposphere controls the climate and ultimately determines the quality of life on the earth.

2. The troposphere is important for life on the earth. The layer contains gases which include oxygen which is used for respiration by animals and carbon dioxide which is used by plants in photosynthesis.

The nitrogen found in this layer also provides an inactive environment for many chemical processes to take place. The gases also support many important chemical processes such as combustion, weathering and oxidation.

3. The “Stratosphere prevents harmful ultraviolet radiation from reaching the earth.

4.The mesosphere, thermosphere and exosphere also prevent harmful radiation such as cosmic rays from reaching the earth's surface.

5.Communication is also made possible by some layers of the atmosphere, specifically the ionosphere.

Global warming

Global warming is the increase of the average temperatures near or on the surface of the earth as a result of what is known as the greenhouse effect. The effect is caused by greenhouse gases. These gases are produced from natural and industrial processes.

The greenhouse effect

1. The greenhouse effect is the process in which the emission of radiation by the atmosphere warms the earth's surface.

   When heat from the sun reaches the earth's surface in form of sunlight, some of it is absorbed by the earth. The rest is radiated back to the atmosphere at a longer wavelength than the incoming sunlight. Some of these longer wavelengths are absorbed by greenhouse gases in the atmosphere before they are lost to space. The absorption of this long-wave radiant energy warms the atmosphere. The greenhouse gases act like a mirror, reflecting back to the earth some of the heat energy which would otherwise be lost to space.

   Sources of greenhouse effect

   The main greenhouse gases are carbon dioxide, methane, chlorofluorocarbons and Dinitrogen oxide.

   •Carbon dioxide

   Carbon dioxide is the main greenhouse gas. The gas contributes over 50% of the greenhouse effect. The following are some of the sources of carbon dioxide in the atmosphere:

   Clearing and burning of vegetation-Green plants, especially woody plants, absorb carbon dioxide from the atmosphere as they grow. When the wood dies, the carbon dioxide is released back into the atmosphere. Clearing of forests (deforestation) and burning vegetation results in the release of carbon dioxide to the atmosphere. The loss of the forests also means that there are fewer trees to absorb carbon dioxide

Deforestation contribute global warming

Deforestation remains high in the world.

•Burning of fossil fuels

Carbon dioxide is a by-product in the combustion of fossil fuels such as coal and petroleum. These fossil fuels are burnt in cars, power stations and industries.

•Methane

The main source of methane is- agricultural activities. It is released from wetlands, such as rice fields and from animals, particularly cud-chewing animals like cows. The emission of methane gas, therefore, increases with, increase in agricultural activities.

Methane is also produced during the mining of coal and oil and when vegetation is burnt.

Since the 1960s the amount of methane in the air has increased by 1% per year, twice as fast as the build-up of carbon dioxide.

Methane molecules have a lifetime of 10 years in the atmosphere.

•Dinitrogen oxide

Dinitrogen oxide is produced from both natural and human-made processes. Human activities which produce, Dinitrogen oxide include combustion of fossil fuels in vehicles and power stations, use of nitrogenous fertilizers, and the burning of vegetation and animal waste.

•Chlorofluorocarbons

Chlorofluorocarbons (CFCs) are organic compounds made up of chlorine, fluorine and carbon. The sources of CFCs in the atmosphere include fridges, air conditioners and aerosols. CFCs are extremely effective greenhouse gases. A CFC molecule is 10 000 times more effective in trapping heat than a carbon dioxide molecule.

Effects of the global warming

The following are some of the effects of global warming:

•Increase in the temperature of the oceans

This causes the bleaching of corals. Bleaching of coral reefs is the loss of pigments and microscopic plant cells from coral tissues. This results in the whitening of the coral reefs.

•Rise in sea levels

Sea levels are rising due to thermal expansion of the oceans and the melting of land ice. This may eventually lead to flooding of coastal lands.

•Change in world's climatic patterns

The climatic patterns in most parts of the world have changed. It is becoming hard to forecast the weather accurately.

Rain no longer falls when expected. Sometimes the rains are heavier than expected, leading to flooding. Other times, the rains are far less than expected, leading to drought.

The extent of the earth's surface under desert condition is also increasing.

•Acidification of the oceans

The world's oceans soak up much of the carbon dioxide produced by living organisms either in. the form of dissolved gas, or from the skeletons of tiny marine creatures that fall to the bottom to become chalk or limestone. '

The carbon dioxide dissolves in the water and forms a weak carbonic acid, thereby lowering the pH of the ocean waters.

Increased acidity and temperatures of ocean waters eventually lead to the bleaching and death of coral reefs.

•Extreme weather events

These include floods, droughts,' heal waves, hurricanes and tornadoes.

Hurricane winds

Other effects of global warming include:

- higher or lower agricultural yields,
- Melting of Arctic ice and snow caps. This causes landslides, flash floods and glacial lake overflow,
- extinction of some animal and plant species,
- Increase in the range of disease vectors, that is, organisms that transmit diseases

Scientists predict that the earth's average temperature will increase by between 1.4 and 5.8°C by the year 2100.

They also forecast that the sea level will rise by at least 25 m, leading to coastal flooding that will displace millions of people. Small islands (such as Zanzibar and Pemba) and low-lying areas will be totally covered by ocean waters.

About 25% of all plant and animal species will become extinct.

Solutions to global warming

The effects of the greenhouse gases in the atmosphere will continue to be felt for a long time. This is because the greenhouse gases remain in the atmosphere for long periods of time. For example, carbon dioxide molecules have a lifetime of between 5O and -100- years in -the atmosphere, while that of CFC molecules is about 110 years. This means that global warming will continue even if we were to cut down on the emission of the greenhouse gases.

However, it is important to reduce on the amount of the emissions before they reach alarming levels. The following are some of the measures that can be taken to reduce greenhouse gas emissions into the atmosphere:

Put in place energy-conservation measures to reduce the use of fossil fuels. These measures include use of public transport to minimize the number of vehicles on the roads and the use of fuel-efficient cars.

Use of cleaner alternative sources of energy, such as solar and wind.

Check deforestation and replant trees (afforestation) that would absorb carbon dioxide.

Afforestation

Countries should commit themselves to minimizing the emission of greenhouse gases into the atmosphere. This is being done through agreements such as the Kyoto Protocol.

Chapter summary

1. Geophysics is the branch of science concerned with the physical, chemical, geological, astronomical and other characteristic properties of the earth.
2. The interior of the earth is composed of three main layers, namely the crust, mantle, outer core and the inner core.
3. The earth's crust and part of the mantle are cracked into huge pieces called tectonic plates. The tectonic plates move about at very low speeds.
4. The line where two tectonic plates meet is called a boundary.
5. There are three main types of boundaries.These are destructive boundaries, constructive boundaries and conservative boundaries.
6. Volcanoes mainly form along destructive and constructive plate boundaries.
7. Hot-spot volcanoes form away from tectonic plate boundaries
8. There are two main types of volcanoes, namely central and fissure volcanoes.
9. Volcanic eruptions affect the landscape, vegetation, wildlife, environment, human life and property, soil formation and the availability of minerals on the earth's surface.

10. Earthquakes occur when rocks within the earth's crust move suddenly, shaking the earth.
11. Earthquakes occur on or near destructive plate boundaries
12. The point within the earth's surface where an earthquake begins is called the hypocentre or the focus of the earthquake.
13. Earthquakes release their energy in three waves called seismic waves. These are primary waves, secondary waves and surface waves.
14. Earthquakes are usually described in terms of their magnitude and intensity.
15. The magnitude of an earthquake is usually measured, on the Richter scale. The scale is .based on the amplitude of the largest seismic wave recorded, no matter what type of wave was the strongest.
16. The intensity of an earthquake is a measure of its strength based on the changes it causes to the landscape. The intensity of an earthquake is measured on the Modified Mercalli scale.
17. A seismograph is an instrument used to record ground-movements caused by an earthquake.
18. There are many hazards associated with earthquakes. They include landslides, tsunamis, collapsing buildings, fire outbreaks and backward rivers.
19. Earthquakes are usually preceded by a number of signs. They include temperature changes, rise or fall of water levels in wells, animals becoming highly disturbed and restless, and an increase in the number of patients and child deliveries in hospitals.
20. The earth's atmosphere is divided into several vertical layers. They are the troposphere, the stratosphere, the mesosphere, the thermosphere and the exosphere.
21. Global warming is the increase in the average temperature on or near the earth's surface as a result of the greenhouse effect.
22. The greenhouse effect is caused by greenhouse gases, which include carbon dioxide, methane, dinitrogen oxide, and chlorofluorocarbons (CFCs).
23. Global warming results in increase in the temperature of oceans, rise in sea level, change in the earth’s climatic patterns, acidification of the oceans, and extreme weather events such as heat waves, hurricanes and drought.
24. Some of the measures that should be put in-place to check global warming include implementation of energy conservation measures,

use of dean alternative sources of energy, checking deforestation, and minimizing emission of greenhouse gases.
