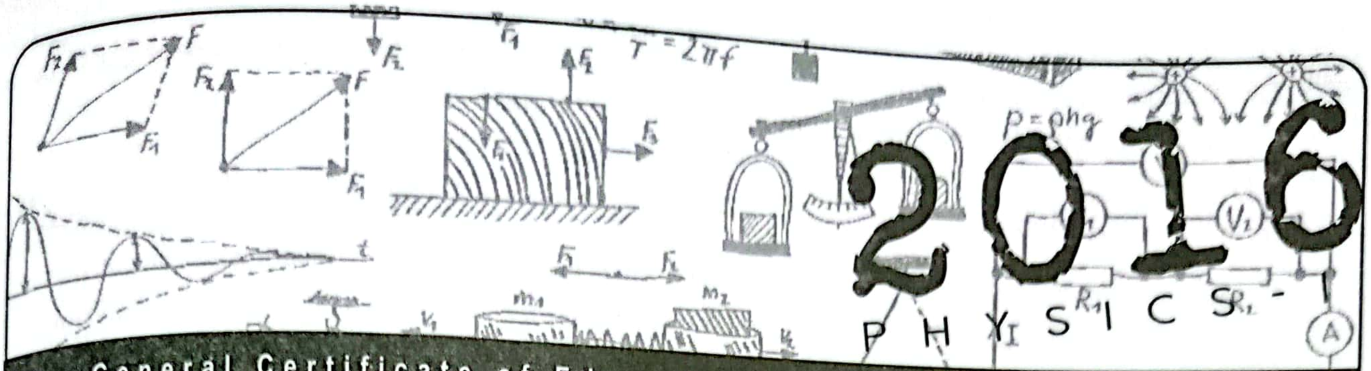


General Certificate of Education (Adv. Level) Examination

01. The SI unit used to measure the activity of a radioactive source is

(1) Bq (2) Gy (3). 1 Be (4) Bq-1 (5) Sv

As soon as you see the question, you can decide that the correct answer is Bq. A similar type of question has been given in 2009 as the first question. A note on Henry Becquerel and the discovery of the radioactivity has been mentioned in my matter and radioactivity book. From the radiation he developed, the photo film which was darkened has been shown here. The bottom figure shows a shadow of a metal cross with darkened parts as the rest. Becquerel might have kept the cross in between the photo film and the sample of Uranium. The alpha particles emanating from radioactive Uranium does not penetrate across metals. Therefore, the location of the cross in the photo film is not exposed to alpha particles.

Such a cross was there in the cathode ray tube of J. J. Thompson. The cathode rays or the electrons did not go across the cross. Sometimes the scientists may have chosen the cross as it is a famous symbol which is made from metals. In addition, the symbol of the cross has a characteristic pattern. As I always tell, many such new discoveries are found by the scientist by a chance. They can be interpreted as the presents of destiny but even the destiny is treating the people who work hard and think something in an innovative way.

Gy is the SI unit of the ionization radioactive dose. If 1 J amount of radioactive energy is absorbed by 1 kg, then the dose of those radiation is interpreted as 1 Gy. Actually, Gy is a derived unit. 1 Gy = 1 Jkg^{-1}. It has been named as Gy to commemorate the British scientist Louis Gray. He was a pioneer who studied the effects of X rays on live tissues.

Sv means the SI unit that measures the health effect from the ionization radioactive dose. Actually, its basic unit is also Jkg^{-1}. There is no talk about the risk to a live tissue form the dose which is measured from Gy. 1 kg can be of anything. When obtaining, Sv from Gy, the amount of Gy is multiplied by a quantity which does not have a unit. Sv has been named to commemorate the Swedish national medical physicist called Sievert.

02. **The percentage error of a certain length measurement has to be kept below 1%. If the error due to the measuring instrument is 1 mm, the measuring length has to be greater than**

(1) 1 mm (2) 1 cm (3) 10 cm (4) 1 m (5) 10 m

Measuring Instruments 01

Cannot you do this question from your memory? If 1 mm is the error from the measuring instrument, then is not it the measuring length be 100 mm to get the percentage error as 1%? Do you have to write expressions? Does not 100 mm mean 10 cm? if you write expressions, then

$1/l \times 100 < 1 \rightarrow l > 100$ mm. The answer is (3). The term 'it should be increased' in the answers is very important. Even it is greater than 1 m, the percentage error can be kept less than 1%. But if we take 1m as the correct answer, then the lengths that are greater than 10 cm and less than 1 m are neglected. Therefore, it should be greater than 10 cm sentence is needed to be taken as the answer. Then all the lengths that are greater than 10 cm are included.

03. **A certain liquid-in-glass thermometer with a uniform bore radius has been calibrated using the boiling point of water and the melting point of ice. Of the following properties, what is the most essential property that a thermometric liquid used in this thermometer must possess?**

(1) high volume expansivity (2) uniform volume expansion
(3) high thermal conductivity (4) low specific heat capacity
(5) low vapour pressure

Thermometry 04

There was an uncertainity in this question. The answer that the examiners expected was uniform volume expansion. But many teachers expressed their arguments to this question by showing relevancy to the 6th question of paper 1994. Here (3) was taken as the correct answer.

In a thermometer

1) The thermometric material should be a liquid through out the total temepraturе range
2) The thermometric material should have a linear characteristic that is varying with the temperature
3) The thermometric material should have a characteristic that changes with the temperature
4) The thermometric material should obey Boyle's law
5) The thermometric material should have a constant expansivity

According to this question this choice is correct. If there is no characteristic that varies with the temperature, then a characteristic that changes in a linear/unifrom way will be just cancel off. If there is no characteristic that does not change with the temperature, then there is no characteristic that changes uniformly. The question of paper 1994 has been asked as a common question. The stem of the question has 'in a thermometer' term only. There is no mention of the callibration or the accuracy or the sensitivity.

In the question of 2016, the thermometer is being callibrated by using the boiling point of water and the melting point of ice. If these two values are only being used, then to callibrate the thermometer correctly, the volume expansivitiy of the liquid should be uniform. There are two stationary points to callibrate. To divide the range between the two values into equal parts, the volume expansivity of the liquid should be uniform. Even it is given that the radius of the hole is uniform. That hint is directing towards the uniform volume expansivity. Why? If the radius is not unifrom, then you cannot callibrate correctly.

Actually higher heat conductivity, lesser specific heat capacity and lesser vapour pressure are the characteristics that a thermometer should have. These are also essential characterisitics. There is an ambiguity in such questions. But the question is directed to a thermometer with a uniform radius in the hole which is callibrated using two stationary points. It is true that the most essential characterisitic is the unifrom volume expansivity of the liquid.

If a material without less vapour pressure like alcohol is being used, then the thermometer cannot be callibrated using the boiling point of water. If it is liquid alcohol, then before 100°C all the alcohol will be evaporated. If so, one can argue that what is the use of unifrom volume expansivity. As the boiling point of water is being used, less vapour pressure can be thought as the essential characterisitc that the liquid have. <u>But the question has mentioned that the liquid-glass thermometer is being callibrated using the boiling point of water and the melting point of ice</u>. Therefore, it is not fair to think that the liquid will be evaporated at the boiling point of water. If the liquid is evaporated, then the thermometer cannot be callibrated using the boiling point of the water.

This question should be answered using the logic. There is a thermometer with a uniform radius of the hole. To callibrate it, the melting point of ice and the boiling point of water are being used. What is the essential characteristic that the liquid should only have for this purpose? Do not think of anything else. Agure according to the given data. Questions like whether the liquid absorbes more heat or does it has a higher heat conductivity or whether the liquid is evaporated on the way are not relevant to the question.

However, in the history of finding answers to the questions of thermometers and thermometric characteristics I have told in many occasions that consider the given facts of the question or else the stem of the question. The nearest example is the 34th question of paper 2012 (new). Here the statement of (A) is true as a sentence. But it does not have a direct effect in measuring the correct value from the thermometer.

Look at the following statements which have mentioned about the ability of a thermometer to give a correct value for a thermometric measurement.

(A) *When the temperatures have to be measured at the instances where it varies quickly with the time, then the given thermometer should have its thermometric characteristic which varies greatly with the temperature.*

(B) *When comparing the heat capacity of the environment that the temperature should be measured, the heat capcity of the thermometer should be negligible.*

(C) *There should be a linear variation in the thermometric characterisitc with the temperature.*

Out of these statements,

1) Only B is true
2) Only A and B are ture
3) Only B and C are true
4) Only A and C are true
5) A, B and C all are true

Therefore, according to the stem of the question, statement A is not matching. Even statement C was taken as the correct answer in 2016, for the question of 2012, there is no relevancy of that statement to the question. The question of 2012 does not ask about calibration. Then you will ask how the thermometer of the question in 2012 was calibrated. If we consider that the thermometer of 2012 as liquid-glass and if it was calibrated using the melting point of ice and the boiling point of water, then definitely statement C is very important. What is your answer if such a question was asked from you? What is <u>the essential</u> characteristic that you should have to pass the exam?

1) Being honest 2) Being humble 3) Not being proud 4) being simple

5) having practiced the subject matter well

All these are characteristics that you should have. There is no debate about it. But the relevant answer according the question is (5). Is not it?

04. Which of the following is not true regarding electromagnetic waves?

(1) Directions of electric and magnetic fields are perpendicular to each other.

(2) Speed does not depend on the medium of propagation.

(3) Do not necessarily require a material medium for propagation.

(4) Direction of propagation of the wave is perpendicular to the directions of electric and magnetic fields.

(5) Can be reflected at the boundary between two media.

When reading the sentences, you will get that the correct answer is (2). The sentences that are in this question have been tested more or less previously. Look at the 13th question of paper 1994 and the 2nd question of paper 2004.

Look a the following statements about electromagentic waves.

A) They have the same speed in any medium.

B) They are transverse waves.

C) There is no need of a medium for their propagation.

Out of these statements

1) Only B is true 2) Only B and C are ture 3) Only A and C are true

4) Only A and B are true 5) A, B and C all are true

Look at the following statements about the plane electromagnetic waves which are propagated in a vacuum.

1) Electromagnetic waves are transverse waves.

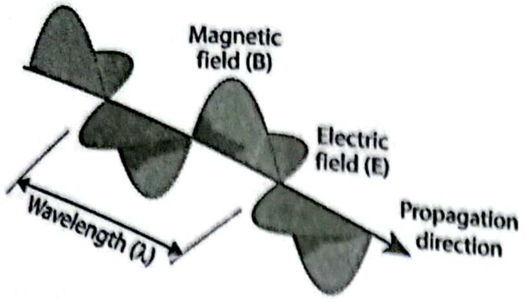
2) The speeds of the electromagnetic waves are independent from their wavelengths.

3) The associated electric and the magnetic fields with the wave are always along the direction of propagation.

Out of these statements

1) Only A is true 2) Only A and B are ture 3) Only A and C are true

4) Only B and C are true 5) A, B and C all are true



The figure has shown the nature of an electromagnetic wave. Here it can be seen how the electric and the magnetic fields are being varied. Always the electric and the magnetic fields are perpendicular to each other. Here it has shown a polarized electromagnetic wave. That means the planes of the electric and the magnetic fields are not changed with the time. In the unpolarized light, the planes of E and B are changed with the time. The normal light is unpolarized. Even the directions of E and B are changed, they are always perpendicular to each other. We cannot see the electric fields or the magnetic fields.

As there is an expression for the speed of any propagating wave, there is an equation for the speed (c) of an electromagnetic wave. This is out of the syllabus. At free space, this speed is given by $c = \frac{1}{\sqrt{\mu_0 \epsilon_0}}$. Here ε_0 is the permittivity and μ_0 is the permeability of free space. Even for a vacuum, there are values for ε_0 and μ_0. Actually, the free space is the vacuum. Therefore, the electromagnetic waves can exist in a vacuum. The electric and magnetic fields can exist in a vacuum.

Mechanical waves are not existing in a vacuum. They need a medium for propagation. Even an electromagnetic wave does not need a medium for propagation, it can exist in a medium. For example, the light is travelled through the glass. For glass, ε is not ε_0. The value of ε is bit higher than $\varepsilon_0$for the glass. Glass does not have magnetic properties. Therefore, for glass μ is μ_0. When ε_0 becomes ε, the speed is reduced. When it goes back to the air, it gets ε_0. Therefore, the speed of light is reduced in the glass whereas it takes the normal value when it comes to the air again. Therefore, electromagnetic waves (light) are dependent upon the speed of the medium. It is wrong to say that it is not dependent. The propagation direction of any mechanical wave is perpendicular to the direction of the molecules' vibration. The propagation direction of an electromagnetic wave is perpendicular to the direction of the fields.

The speed of electromagnetic waves that are propagating in the vacuum is independent from their wavelengths. If it is in a medium, then the speed is dependent upon their wavelengths. The speed of red light is not equal to the speed of blue light in the glass. But at vacuum, all the colours have equal speeds.

When you think of an electromagnetic wave, think of the light. The light can be subjected to reflection at the interface of two mediums.

05. A student has suggested the following three methods (A), (B) and (C) to increase the voltage sensitivity (V/cm) of a potentiometer wire.

(A) Increasing the length of the wire
(B) Connecting a resistor in series with the wire
(C) Increasing the voltage applied across the wire

Of the above three methods,

(1) only A is correct.
(2) only A and B are correct.
(3) only B and C are correct.
(4) only A and C are correct.
(5) all A, B and C are correct.

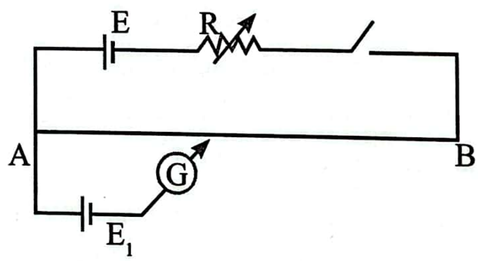
Potentiometer 08

This question faced a controversial situation. Is there any definite interpretation for the voltage sensitivity of a potentiometer wire? Is it cm/V or V/cm? The current sensitivity of a galvanometer is the deflection (θ) per a unit current (I). That means θ/I. The sensitivity of the glass-liquid theremometer is the rise of the liquid level for a certain temperature difference. If the sensitivity is high, then it shows a high expansion length for a particular temperature difference. Then you can even measure a small temperature difference conveniently.

The sensitivity related to getting angry is measured by the amount s/he gets angry/ the reason to get angry. Does the voltage sensitivity written as V/cm in the brackets just after the term? Actually, V/cm means the potential drop per a unit length. That menas the potential gradient. The potential gradient should be less if the voltage sensitivity has to be increased. There is no argument in it. The term given in brackets can be the potential gradient not the sensitivity.

Children know the voltage sensitivity. This has been checked before. As V/cm is given in bracekts, one may think that V should be increased to increase the sensitivity. There is no such answer relevant to that. If the voltage sensitivity is interpreted like cm/V, then you can just get the answer. If V is kept constant and the length of the wire is increased, then the sensitivity is increased. Then the potential gradient k =V/cm will be decreased. To increase the sensitivity you need to reduce k.

When a resistance in series is connected to a wire, the potential drop across the wire gets reduced.Then k gets reduced and the sensitivity gets increased. When the potential drop is reduced across the wire, you need to go for a longer length to measure a certain potential difference. That means the potentiometer is sensitive. When the voltage is increased across the wire, then k will be increased. The sensitivity is reduced. Therefore, what is more correct will be (A) and (B) only. Look at this question which is given as the 26th question in paper 1991. You can do a study in detail once you do this question.



To decide the e. m. f value of the cell E_1, there is a potentiometer method arrangement which is shown in the figure. To increase the balanced length,

1) R should be decreased and E should be increased
2) R should be decreased and E should be kept constant
3) E should be increased and R should be kept constant
4) R should be increased and E should be kept constant
5) The diameter of the potentiometer wire should be decreased

When R is decreased and E is increased, V (the potential drop across the potentiometer wire) will be increased and the sensitivity is reduced. Argue and see. The correct answer is (4). When the diameter of the wire is decreased, the resistance of the wire increases. Then the current will be reduced and the potential drop across the wire (E -iR) is increased. V is increased and the sensitivity is reduced.

06. In a certain transformer there are 360 turns in the primary coil and 30 turns in the secondary coil. Which of the following voltage conversions is done using this transformer?
(AC = Alternating current, DC = Direct current)

(1) 240 V AC voltage to 12 V DC voltage.
(2) 240 V AC voltage to 2 880 V AC voltage.
(3) 240 V DC voltage to 20 V DC voltage.
(4) 240 V AC voltage to 20 V AC voltage.
(5) 240 V DC voltage to 2 880 V DC voltage.

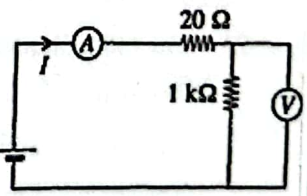
Electric Field 10

It is very simple. You just do not have to look at the sentences with direct currents. Such questions are given in the previous papers. When direct currents are removed, only (2) and (4) are left out. The number of turns are greater in the primary than in the secondary. Therefore, this is step-down transformer. That means the secondary voltage should be lesser than 240 V. Only 4th choice will be left out. The number of turns are reduced from 360 to 30. That means by 12 times the turns have been reduced. When 240 is divided by 12, you will get 20. Do not do rough work. Do you need rough work to divide 360 by 30? Even do you need rough work to divide 240 by 12? Even if you do not do the calulation, from the relevant logic you can get the 4th choice as the answer. The statements with direct currents should be removed. Cut off the answers that have more than 240 V for the secondary voltage. When 240 V is multiplied by 12, you will get 2880 V. Look at the 5th question of paper 2009. 220V has been changed into 240 V this time.

To change 220 V ac to 20 V ac, which of the following transformer is suitable?

	Ransformer Type	*No. of turns in the secondary coil/ No. of tunes in the primary coil*
1	Step-down	*1/22*
2	Step-down	*1/11*
3	Step-down	*11*
4	Step-up	*1/11*
5	Step-up	*11*

07. Of the following sets of internal resistances given, the set of internal resistances that suits best for an ammeter (A) and a voltmeter (V) to have in order to measure the current I and the voltage across 1 kg resistor of the circuit shown is



	Internal resistance of ammeter	Internal resistance of voltmeter
(1)	1Ω	5kg
(2)	5Ω	1kg
(3)	1Ω	20Ω
(4)	20Ω	5kg
(5)	5Ω	50kg

Moving Coil Meters 08

This is also very simple. In an ammeter, the internal resistance should be small and in a voltmeter, the internal resistance should be large. This is a common generally fair statement that we express. Even if you look from this point, the internal resistance is minimal in 1 Ω for the ammeter out of the given numbers.

For the internal resistance of the voltmeter, the biggest value from the given values is 5 kΩ. Therefore, the correct answer is (1).

The voltmeter is connected parallel to 1 kΩ. Therefore, the internal resitance of the voltmeter should be definitely greater than 1 kΩ. From the given values, 5 kΩ is the only value which is greater than 1 kΩ. Therefore, all you need to look at only (1) and (4). In ammeter, it is better to have a lower resistance for its internal resistance. Then the change of the equivalent resistance of the circuit is negilgible even if we put the ammeter. If we think that voltmeter is ideal without the ammeter, then the equivalent resistance of this circuit is 1020 Ω. When an ammeter with an internal resistance of 1 Ω is connected,then the total resistance will be 1021 Ω. The change is very small.

The percentage change = (1/1020) X 100% ≈ 0.1%

When an ammeter with an internal resistance of 20 Ω is connected, then the total resistance is increased upto 1040 Ω. The percentage change = (20/1020) X 100% ≈2%

The change of 0.1% is smaller than 2%. It is better to reduce the errors as far as possible. It is better if the internal resistance of the voltmeter is increased more. When 1 kΩ and 5 kΩ are parallel, then the equivalent resistance is 0.83 kΩ. Then the percentage of change = (1-0.83)/1 X 100% ≈ 17 %

But from the given internal resistances, you can pick only 5 kΩ for the voltmeter.

8. Which of the following is not a result of surface tension?

(1) Formation of spherical water droplets

(2) Capillary rise of water

(3) Ability of insects to walk on water surfaces without sinking

(4) The excess pressure inside a soap bubble

(5) Escaping of water molecules from water surfaces

Surface Tension 10

It is very simple. You know these things. The water droplets are spherical in shape due to the forces of surface tension. Look at the 6th question of paper 1987.

What happens due to the surface tension from the following effects?

1. The rising of a mercury column of a thermometer when the temperature is increased
2. The falling liquid drops takes a shperical shape
3. The rising of a mercury column in a barometer when the atmospheric pressure is increased
4. Acquiring a uniform velocity by a spheical object that falls down in a fluid
5. The spread of fragrance in the room when the scent bottle is opened

We know that for a given volume, the minimum surface area can be obtained for a sphere. The surface energy is proportional on the area. Therefore, acquiring a minimal area is connected to minimum surface energy. Then the potential enrgy of the water droplet (from cohesive forces) will get minimum. The nature always tends to keep its possessed energy at a minimum level. When the energy is at a minimum, there are less troubels. There is no harm for others too. Due to gravitational forces, the water drops do not take a spherical shape. Water droplets are mostly spherical in shape. The 3rd esay question of paper 2010 is about this fact. Normally we consider as a water droplet when the radius of it is less than 0.1 mm. We know that capillary rise, the ability of walking on the water surface by some insects (22nd question of paper 2015) and the extra pressure inside the soap bubble are the results of surface tension.

The emission of water molecules from a water surface is explained by the kinetic theory. The molecules with higher kinetic energy have a great probability to go out from the surface. This is known as vapourization. We cannot say that surface tension forces are not affecting the removal of liquid molecules from the liquid surface. For example, the liquids that have less surface tension forces are quickly vapourized whereas they can be kept as a sytem with very small particles in aerosol form.

But the emission of a water molecules from a water surface should not be taken as a result of surface tension is not an issue is my view. But, the surface tension affects to emit or not.

9. **Consider the following statements made about a standing wave on a stretched string.**

(A) The energy does not propagate along the string.

(B) The position of a node does not vary with time.

(C) Maximum displacement achieved by each particle in the string depends on its position along the string.

Of the above statements,

(1) only A is true. (2) only B is true. (3) only A and C are true.

(4) only B and C are true. (5) all A, B and C are true.

Transverse Waves **03**

These sentences are known by yourself. Look at the 12th question of paper 1990 and 23rd question of paper 1991.

Out of the sentences below, which is false regarding the standing waves?

1) The wave model does not go with the standing wave.

2) There is no energy propagation along with the wave.

3) You need two waves for wave superposition and they can travel to the same direction or to the opposite direction.

4) In wave superposition, always there will be points with zero displacements.

5) The points of maximum displacement are situated in the middle of zero displacement points.

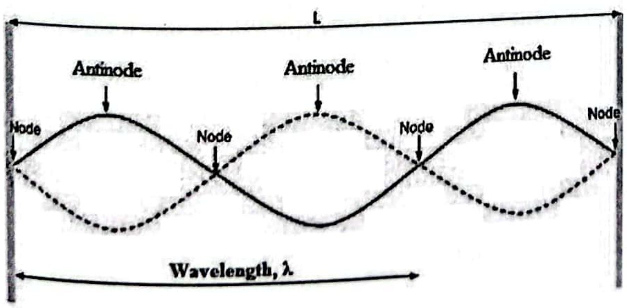
Look at the following statements about progressive waves and standing waves. Which is in false?

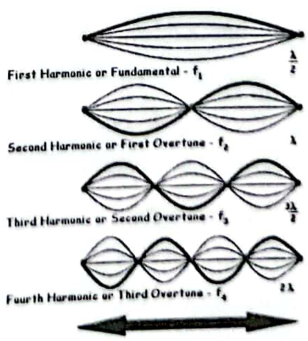
A) In a progressive wave, each particle of the string will vibrate with the same amplitude.

B) In a standing wave, each particle of the string will vibrate with the same frequency.

C) The amplitude differes for different particles in the standing wave.

From these two questions, you can find the correct answer for the question of 2016. The standing wave patterns in a stretched string are shown in the figure.





The standing waves are created by the superposition of two waves which travel in opposite directions in a medium with same wavelength or frequency. The wave model of a standing waves created from this way does not go to the left or right. It stays on the same place. The term standing has been given to it due to this reason. At some places of the string (nodes) are at rest every time. In the middle of two adjacent nodes, there is point which has the maximum displacement (antinodes).

In the question of 2016, all three sentences are correct. There is a tension in the string. Therefore, there is elastic energy in the string. But there is no energy propagation through the string. The location of nodes and antinodes do not change with the time. They are stable with the time. By looking at the wave model you can say that the statement (C) is correct. At the nodes, the displacement is minimum or zero. The maximum displacement of each particle is characterisitic to each particle.

Can you call the maximum displacement of each particles as their amplitudes? There is no wrong in that. But in a standing wave, there is no such thing as one and only amplitude. It is dependent upon the location of the particle. The maximum displacement of a standing wave (maximum amplitude) is at antinode. In a progressive wave, at a certain point of each particle, will acquire the maximum displacement or the maximum amplitude. Each particle's maximum displacement differes in a standing wave. Those values are not being changed with time (if there is no energy loss).

You can understand these facts clearly through mathematics. If you are a mathmatics lover, then read these information. A sinusoidal progressive transverse wave which travel towards positive x direction can be represented mathematically like this. $y_1 = A\sin(kx - \omega t)$

Here x = the position of the wave in time t; A = the amplitude of the wave; $k = 2\pi/\lambda$ = the wave number of the wave (λ = wavelength); ω = the angular frequency of the wave; y_1 = the displacement of the wave at a certain (x, t)

Another similar wave which travels in -x direction can be represented by $.y_2 = A\sin(kx + \omega t)$.

To get a standing wave, we need to superpose these two waves.

Here $y = y_1 + y_2 = A\sin(kx-\omega t) + A\sin(kx+\omega t)$

When the above expression is simplified [according to sin A + sin B = 2 sin ½ (A + B) cos ½ (A-B)], $y=[2A\sin(kx)]\cos(\omega t)$.

Here from the term $[2A\sin(kx)]$ gives the maximum displacement of a certain x position (it gives the amplitude relavent to that point). From the term the oscillation of the standing wave with time is represented (the up and down of a certain x point). A is the amplitude of the progressive wave. It is a constant for a progressive wave. The term $2A\sin(kx)$, does not interpret the amplitude of the standing wave any time. Because, the term $2A\sin(kx)$, is dependent upon x. Actually, there is no characteristic amplitude to a standing wave.

When considering the above expression, the term which oscillate with time does not have x. It has only $\cos(\omega t)$. What is implied here is that there is no progression in the oscillating part. That means the wave is stable relative to x. The above expression which represents a standing wave tell us that, a certain point x moves up and down and there is no motion to the left or right. When $kx = n\pi$ (n = 0, 1, 2, …), then the term will be zero. This is because $\sin 0 = \sin\pi = \sin 2\pi = 0$.

The term $2A\sin(kx)$ will be zero for the values of x when $kx = n\pi$.

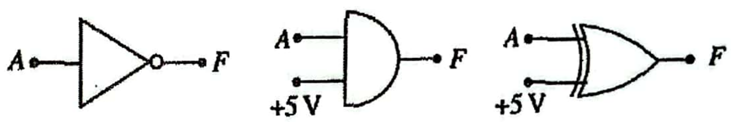
$2\pi/\lambda.\ x = n\pi \rightarrow x = n\lambda/2$ (n = 0, 1, 2, …). Do not these x values give nodes? Likewise, the term will be maximum when $kx = 1/2\,\pi, 3/2\pi, 5/2\pi, \ldots$. That means $kx = (n + ½)\pi$ (n = 0, 1, 2, …). From this the position of antinodes will be obtained as $x = (n + ½)\lambda/2$.

From these expressions, we can easily decise that the distance between two adjacent antinodes is $\lambda/2$ and they are placed in between the middle of a pair of nodes.The maximum displacement of an antinode (amplitude of an antinode) is 2A. When both of them are properly joined what else should happen other than 2A? Therefore, the maximum displacement (amplitude) of the particles of a standing wave is ranging from a minimum to 2A. A is the ampltude of the waves which has participated to make a standing wave. Another characteristic of the equation which represents a standing wave is that there is same ω for any value of x. Therefore, each particle vibrates with the same frequency. The speed of the particles with maximum displacement is greater compared to the speed of particels without maximum displacement. It should be like that. If the frequency is same, then it should go quickly if it needs to go a long distance. You do not have to go quickly to cover a small distance.

Here I will mention another question which is asked by the teachers from me. Does standing waves is a result of wave interference? My answer is this one. When two waves are going in the same direction, the result will be an interference. To have a standing wave, you definitely need two waves that moves to the opposite direction. But the two effects occur due to the wave superposition principle. We use interference term to describe the results of the superposition of waves to the same direction. The superposition principle is a very powerful principle in Physics. Even it is seen as a simple principle, if it does not occur we will undoubtedly be desperate. This principle can be simply state like this way. When many effects are being acted simultaneously, the net resultant is equal to the vector sum of the results of each individual effect. As mentioned in a previous book, this principle cannot be applied to human activities. Human activities are dependent upon the others around them.

10. **Which of the following gates operates according to the truth table given?**

A	F
0	1
1	0



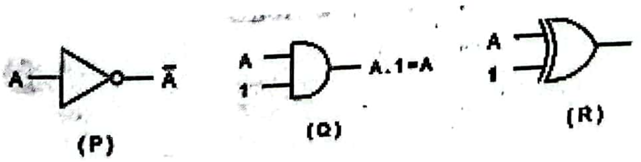
(P) (Q) (R)

(1) P only (2) P and Q only (3) Q and R only

(4) P and R only (5) All P, Q and R

Logic Gates **09**

There is a similar question like this as the 21st question in paper 2003. If you take +5 V as binary 1 and write the output at the end of each gate, then the answer is in your hand.

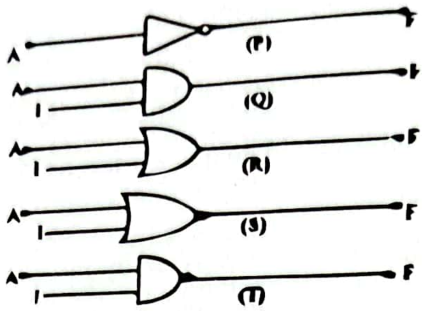


$\bar{A}.1 + A.\bar{1} = \bar{A}\ (\bar{1} = 0)$

From this you can see that (P) and (Q) gates action is identical. It can be directly seen that the given truth table is the truth table of a NOT gate. Therefore, (P) just matches to the relevant truth table. In (Q), if we take +5 V as binary 1 and when A is binary 0, then the output of AND gate F gets zero. If F =0 when A =0, then the given truth table is not being satisfied. AND gate is a marriage gate. The output gets 1 only when both are there.

In (R) there is an exclusive OR gate (XOR). When the two inputs of a XOR gate is same, then the output zero. XOR gate differes from OR gate in the instance where both inputs get binary 1. In a OR gate, the output is 1 when both of the inputs are 1. But in a XOR gate, the output is 0 when both of the inputs are 1. In any other instances the XOR gate is equivalent to the OR gate. Even both are there or not, the output of a XOR gate is 0. It is enough to stay one (exclusive OR).

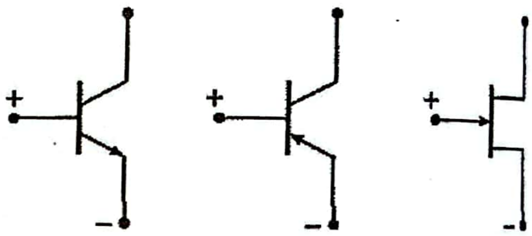
When +5 V is taken as binary 1, then F =1 when A =0 (there is one). When A =1 as both are there, F =0. Therefore, it is equivalent to a NOT gate. XOR gate is like a couple who afraid of each other. Even if both are there or not, there are no shoutings. Even if there is one, then s/he will be the heroine/ hero without the fear.



The second input of each shown gates are being connected to binary 1. Out of the gate which gates are identical in action? (2003)

1) Only P and Q 2) Only Q and R 3) Only R and S
4) Only S and T 5) Only P and T

11. Which of the figures shown correctly indicates the polarities of potential difference that have to be applied across the junctions shown in order to operate the transistor properly and obtain a suitable current?



(1) in A only (2) in B only
(3) in C only (4) in A and C only
(5) in B and C only

Transistors 09

As mentioned in the review of 2014, the arrow of the transistor is always applied from p to n. Therefore, (A) is a npn type bi-polar transistor that you know. To activate a npn transistor, the base-emitter juction should be in the forward biased mode. If it is a Silicon transistor, then V_{BE} should be 0.7 V. (A) is correct. Actually, if (A) is correct, then (B) should be wrong. There is pnp type transistor in (B). The arrow is from p to n. Therefore, the polarities should be to the other side. (C) has shown a n-channel JFET. In n-channel JFET, gate(G)-source(S) junction should always be kept in the reverse biased mode. From that, the electrons emitted from the source can be controlled in going to the drain. If the gate-source junction is made forward biased, then there will not be a depletion region between the gate and the channel. The electrons tend to flow into the gate. Therefore, only (A) is correct. These points (with a relevancy to the transistors) are there in the 28th question of paper 2010 and the 34th question of paper 2011 (new syllabus).

12. When the body temperature of a person is 35°C, the peak wavelength of the radiation emitted from the body occurs at 9.4 p.m. If his body temperature increases to 39°C, the peak wavelength will be (Assume that the black body radiation conditions can be applied)

(1) $\frac{35}{39} \times 9.4\ \mu m$ (2) $\frac{39}{35} \times 9.4\ \mu m$ (3) $\frac{77}{78} \times 9.4\ \mu m$

(4) $\frac{78}{77} \times 9.4\ \mu m$ (5) $\left(\frac{78}{77}\right)^4 \times 9.4\ \mu m$

Radiation **11**

This is a question related to Wien displacement law. As there is no need to simplify till the end, it is easy. When we use $\lambda_m T$ = constant, the answer can be obtained. The given temperatures should be converted to Kelvin units. If you cannot add 273 to 35 by your memory, then write it on the rough paper and add. 35 = 273 = 308. Now, do not try to add 273 to 39 again. 39 is 4 more than 35. So, 308 should be 312. Now the answer is (94 X 308)/312. There is no such answer. It indicates that 308/312 has been simplified. When 308 is divided by 4 it is 77 whereas 312 is divided by 4 it is 78. The answer is (3). If 308 is divided by 4 is 77, then when 312 is divided by 4 it should be 78. As 312 is 4 more than 308. The answers (1) and (2) are in ^{0}C. These cannot have fourth powers. When the temperature is increased, λ_m should be decreased (according to $\lambda_m T$ = constant). So, the correct answer is (3). Is not it?

13. A moving jet plane can create a maximum sound intensity level of 150 dB. Take the sound intensity at the threshold of hearing as 10^{-12} W m^{-2}. The maximum intensity of the sound that can be created by the jet plane in W m^{-2} is

(1) 100 (2) 200 (3) 400

(4) 800 (5) 1 000

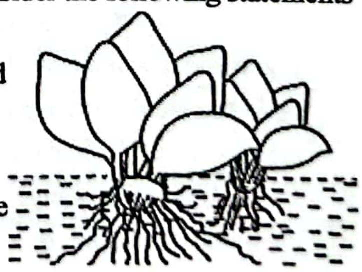
Intensity of Sounds **03**

This also can be done using the memory. As there is 10 in the decibel formula, when you remove a zero in 150, only 15 is left on the left side of the equation. So, in the log expression there should be 10^{15}. Its denominator has 10^{-12}. When it is taken to the numerator, it will be 10^{12}. Then there should be 10^3 to be 10^{15}. The answer is 1000. If you do the calculation, then

$150 = 10 \log I/10^{-12} \rightarrow I/10^{-12} = 10^{15} \rightarrow I = 10^3$

Even in the answer, there should be only powers of 10. If there are other numbers, then you cannot get the answer without looking up the logarithmic tables. Therefore, the correct answer should be 100 or 1000. From 10^{12}, you need 10^3 to take 10^{15}.

14. When wind blows over the surface of a still lake, a bunch of water hyacinth floating on water as shown in figure is observed to move in the direction of the wind with a velocity v. Consider the following statements made about v.



(A) Magnitude of v depends on the rate at which the momentum is transferred from air molecules to the bunch.

(B) Magnitude of v depends on the viscosity of water.

(C) Magnitude of v depends on the mass of the bunch. Of the above statements,

(1) only C is true. (2) only A and B are true. (3) only B and C are true.

(4) only A and C are true. (5) all A, B and C are true

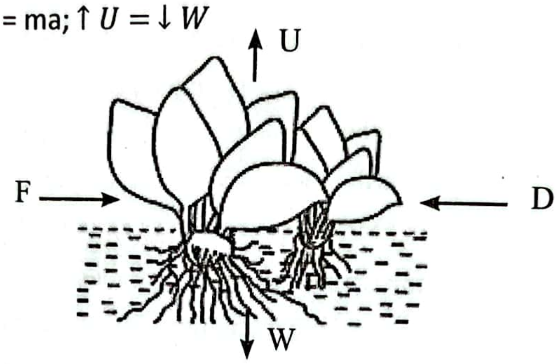
Newton's Law and Momentum **02**

This is not a hard question. Depending upon the rate that gas molecules transfer momentum to the bush, the force that is obtained by the bush changes more or less. If the wind is blown quickly, the rate of momentum transfer to the bush will take a higher value. Then the magnitude of the acceleration of the bush gets greater. Therefore, the magnitude of v is dependent upon the rate in which gas molescules transfer the momentum to the bush. This can be said even from the general knowledge. If the wind is blown slowly, then a particular thing will be pushed slowly in a slow speed. If the wind is blown quickly, then it will be pushed quickly.

When the bush is moved, the visocity force from the water is affecting on the part that is sunk in water. The figure shows the forces that are acting on the moving bush.

W – the weight of the bush; U – the upthrust on the bush; F – The force that is given to the leaves by the wind; D- viscous force

At a certain moment if a is the acceleration of the bush, the for the mass m of the bush we will apply $\rightarrow F = ma$. Then F - D = ma; $\uparrow U = \downarrow W$



D is not a constant eventhough F is considered as a constant. We know that the viscous force is dependent upon the velocity. Therefore, D can be written as D = kv where k is a constant. It is clear that the acceleration of the bush is dependent upon the viscosity of water. So, v is also dependent upon the viscosity of water. Therefore, statement (B) is also true.

There were some arguments for statement (C). Clearly the mass m of the bush is dependent upon the acceleration of the bush. If m is greater, then the acceleration is less. If m is less, then the acceleration will be greater. Therefore, the magnitude of v is dependent upon m. If v is the terminal speed of the bush, then once it attains the terminal velocity the value of v will not be dependent on m. When the speed of the bush gradually increases, the viscous force (kv) also gradually increases. Therefore, initially the speed of the bush will not take the same value. If we consider that F is constant, as v is increased when the value of D is increased, a is gradually reduced. At a certain point when F = D then a = 0. Then from that point, the bush will move at a uniform speed. Once it gets the uniform speed or the terminal speed, that speed is not dependent upon m. It is true.

But the value of the terminal speed (the acquired uniform speed) is dependent upon the mass of the bush. As v can be the speed of an instance, or the terminal speed, the three statements are true for any occasion. Once it gets the terminal speed, it is true that v does not change. But the numerical value of v value is dependent upon the facts that are mentioned in the three sentences. Even if m is greater, then the weight of the bush is increased. As the weight should be equal to the upthrust, if m is greater, then most of the part of the bush should be sunk in water. When the bush moves, the viscous force from the water on the bush also should be increased. So, however v is dependent upon m.

To minimize the viscous force, the net area to the direction of the moving object should be minimized. If most of the part is sunk in water, then the net sunk area of the bush is increased. What is written relevant to the first essay question of paper 2015 is presented here. If an athlete who runs 100 m will press the fingers of the palm and keep it parallel to the direction of motion, then that is to reduce the area. At any time he/she will not keep the palm perpendicular to the moving direction. The participants of swimmimg competitons wear tight swimwear to reduce the viscous force from the water. They remove the hair on the body to get the advantage even from 1/1000 of a second. If they have a long hair, definitely the swimmers tie up their hair and cover the hair with a smooth cap.

Do you get the secret that when the birds fly, the reason why they keep the beak bent downwards and keep their legs up parallel to the body and when an aeroplane is lift off, why the wheels are inserted to body? Horse riders as well as the bicycle riders even participate the competition by bending forward to reducce the frontal area. It is better that trees are bent for a rough wind. Then the forntal area that is associated with wind will be reduced.

15. **An object falling down vertically in air suddenly explodes into four pieces. Which of the following diagrams shows the possible directions of motion of the pieces immediately after the explosion? (↓ - direction of the object before explosion)**

(1) (2) (3) (4) (5)

You can nicely argue and get the answer easily. As the direction of the motion of the object was vertically downwards before the explosion, the momentum of the combined system should also be vertically downwards after the collision (conservation of momentum). Therefore, the net mometum of the exploded parts cannot be in horizontal and as well as in upward direction. Apply this logic to every figure.

Figure 1 - There a piece which goes horizontally to the left side. The cancel off its momentum, there is no piece which goes horizontally to the right side. (1) is wrong.

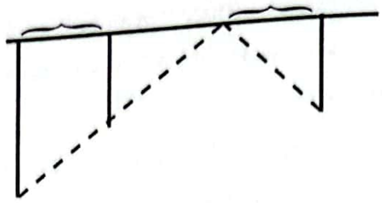
Figure 2 – There is piece which goes horizontally to the right side. To cancel off its momentum, there is no piece which goes horizontally to the left side. (2) is wrong.

Figure 3 – All of the parts go upwards. There is no component which goes vertically downwards. It is wrong.

Figure 4 – There is piece which goes downwards to the right side. The right side of its momentum has a horizontal component. To cancel off, there is no horizontal momentum component piece to the left side. As in figure 2, all of the other parts go vertically upwards or downwards.

What is left will be figure (5). If the the momentum (velocity) of the moving pieces which are inclined are resolved horizontally, then there are horizontal components to the left and right side. Therefore, these can cancel off with each other. They can be equal and opposite too. All the other components are vertically downwards. If it is simply said, then there is nobody to cancel off the momentum or the components of momentum which are to the left or right horizontally in (1), (2) and (4). In (3), all are there to upward direction. What is left will be (5).

The drawn arrows represent the direction of the pieces/parts. They are not the vectors which represent the momentum or velocities. If you think like that, then you will be in trouble. If these are taken as momentum vectors, then once the inclined vectors to the right and left of figure (5) are resolved, they will be cancelled off with each other. Then there is nobody to cancel the horizontal momentum to the right. The arrows only show the directions.



If they are vectors, then once we take the magnitude of the two momentum vectors to the right side in figure (5) as equal, the magnitude of the inclined vector to the left side should be greater than the magnitude of the previous two vectors (look at the figure). Why? Its horizontal vector component of the momentum to the left side should be equal and opposite to the total of the horizontal vector component of the momentum to the right side and the horizontal component of the inclined momentum vector to the right side.

16. The two straight lines shown in the displacement (d) —time (t) graph represent the motions of two objects A and B started from rest at time t = 0 and moving along the positive x-direction. Which of the following statements made about the motions of the objects is true?

(1) The object A has travelled for a longer time than B.

(2) When t_0 object B has made a displacement greater than A.

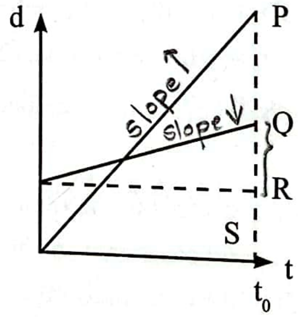
(3) Object A has a greater velocity than B.

(4) Object A has a greater acceleration than B.

(5) Both objects have the same velocity at the point where the two straight lines cross each other.

Linear Motion

It is a very easy question. Both have gone in the same t_0 time. That can be decided from the eyes. (1) is false. Even at a glance we can say that the statement (2) is wrong. The object A has done more displacement than the object (B). Actually, the displacement of B is the distance of QR. A has displaced a distance of PS. PS> QR. If needed PS> QS too. You can quickly see that statement (3) is correct. The gradient of the displacement-time graph gives the velocity of the object. Even a person with eyes can see that the velocity of A is greater than B. The gradient of the straight line of A is greater than of the straight line of B.



Now you do not have to look into other two sentences. As the displacement-time graphs are straight lines, the velocity is uniform. The acceleration is zero. (4) is false. As (3) is correct (5) is obviously wrong. The place where the straight lines intersect the displacement values of the objects are equal relative to the starting point. But really the displacements are not equal. According to the drawn graphs, at t =0, there are velocities to the two objects. Even at t =0, there are gradients for the straight lines. Many teachers said that this question is wrong to me as the question has mentioned that at t =0, the objects start from the rest. But there is no effect to the answer of the question from that.

17. **An elevator of weight 5 000 N carries a load of 5 000 N. While moving vertically upwards in a building, it travels at constant velocity from 2nd floor to 12th floor in 20 seconds. The height of each floor is 4 m. If only 80% of the power generated by the motor is consumed to lift the elevator and the load against gravity while moving at constant velocity, the power of the motor is**

(1) 20 kW (2) 25 kW (3) 40 kW
(4) 60 kW (5) 1 000 kW

Work Power and Energy **02**

If you love my books, then you should first start to do rough work (calculations) from this question. There is a simple calculation. There are 10 floors from the 2nd floor to the 12th floor. The total weight of the lift with the load is 10^4 N (5000 + 5000). The total height of 10 floors is 40 m. Therefore, the work done against the gravity should be (10^4 X 40) J (mgh = wh). Therefore, the work done in a second is (10^4 X 40)/20. If the power of the motor is P, then (10^4 X 40)/20 = P X 80/ 100; P = (2 X 10^4 X 10)/8 = 25 X 10^3 W = 25 kW.

If you think that there are 11 floors from the 2nd floor to the 12th floor, then the calculation gets wrong. From the second-floor means, that the second floor should be included. Up to the 12th floor means, the 12th floor is not included. Up to the 12th floor means the end of the 11th floor. A similar question of a previous paper has been mentioned to imply the importance of studying the past papers in an investigative manner to you.

An electric motor pulls a weight of 100 kg to 20 m in 2s. If the efficiency of the motor is 80%, then what is the minimum power that the motor should have?

18. **Three monochromatic light beams A, B and C have the same intensities (i.e. energy flow through unit area per second). However, the wavelength associated with beam A is longer than that of B, and the frequency associated with beam C is smaller than that of A. The photon flux (number of photons crossing a unit area per second) of three beams when written in the ascending order, it will be**

(1) C, A, B (2) B, A, C (3) A, B, C
(4) B, C, A (5) C, B, A

Radiation **11**

We will consider a group of giants and a group of people like us. To finsh a certain work if you need a certain amount of energy, then you will need few giants. You will need many people like us for that work.

The example has been presented in an identical way which is based for this question. The energy that flows across a unit area in a second is same for all the three rays. But according to the photon theory, the energy of a photon with a less wavelength is greater compared to a photon with a high wavelength ($E = hf = hc/\lambda$). Therefore, if we take as a group and you need to get the equal energy amount, then you need many number of photons with higher wavelengths. As the energy of the photon with less wavelength is high, if we consider as a group, then only few of them are needed to get the equal amount of energy (like the giants). Therefore, to get the correct answer, the relevant wavelengths of A, B and C have to be compared. According to the fist data, $\lambda_A > \lambda_B$. Write down this fact on your rough sheet. From the second data, frequency has been compared $f_C < f_A$. If you try to argue with both frequencies and wavelengths, all will get jumbled. Therefore, the frequency difference has to be changed into wavelengths. If $f_C < f_A$, then $\lambda_C > \lambda_A$.

Now if $\lambda_A > \lambda_B$ and $\lambda_C > \lambda_A$ is expressed as a relation of inequality, then you can write as then $\lambda_C > \lambda_A > \lambda_B$. Now the answer is in your hand. The wavelength is highest in C. Therefore, we need more from it. The wavelength is lesser in B. Therefore, we need few of it. So, when they are made in the ascending order, the correct answer is B, A, C.

You can get the answer by looking at the inequality of λ. You can get confused even it is not a hard question. As two data are given in wavelengths and frequencies, I guess it will be hard if you try to argue separately.

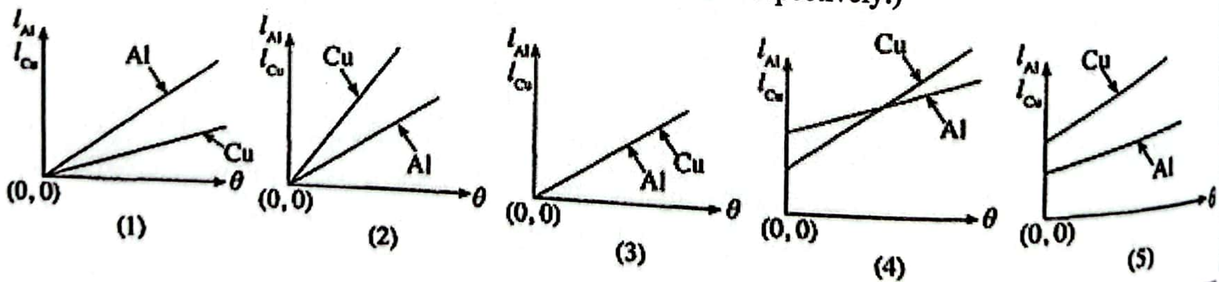
So, here I have built the inequalities using the wavelengths. If needed, you can also do it from the frequencies ($f_C < f_A < f_B$). The intensity and the photon flux density have been given separately in the brackets. I feel it is better to mention like that. In the wave theory, the intensity is interpreted as the energy that flows across a unit area in a unit time (Wm^{-2}). But in the photon theory, we displace waves with the photons. There what is important is the number of photons per unit area in a unit time. This is also known as the photon flux density. As we talk about a unit area, I feel it is better to use the term density. Most of the time, when we talk about photons also we use the intensity of a beam. But here what is indicated as the intensity is the photon density.

According to the photon theory, if the intensity of a beam means the energy that flows across a unit area per second, then there are three ways which we can change the intensity. One way is changing the numebr of photons per second without the change of the frequency (wavelength). Then the energy of each photon does not change but as more or less photons flow across the unit area, the net energy flow per seocnd will be more or less. The second method is to change the frequency (wavelength) without changing the photon flow rate. Then as the energy of the photons will be more/less, the net energy flow across a unit area per second will be more/less even if the rate is constant. The third method is to change both of these facts.

But accroding to the photon theory, the scientists clearly state that the electron emission from a certain surface will not be decided by the intensity of the incident beam. According to the photon theory, if the intensity of the beam is the energy that flow across a unit area per second, then then the intensity can be changed from the second method which was mentioned above. That means the intensity can be changed by changing the frequency (wavelength). What it implies is that when the frequency is increased, the intensity of the beam is increased and the kinetic energy of the electrons gets increased in the emitted electrons due to the intensity (if the frequency increased the threshold frequency).

But what scientists state is that the kinetic enrgy of the emitted electrons will not be changed due to the intensity of the beam. If so, according to the photon theory, the intensity is the number of photons that go across a unit area in a second (photon flux density). By changing the energy of a photon, the photon flux density cannot be changed. If the photon does not have the minimum energy to grab an electron, then there is no use even if many photons are hit. Even if one boy does not have the energy to win the heart of a girl, then there is no use, even many boys are there to win her heart.

19. **l_{Ai} and l_{Cu} respectively represent fractional increase in the original lengths of two rods of aluminium (Al) and copper (Cu) when their temperature is increased by an amount of 0 °C from the room temperature. Which of the following graphs best represents the variations of l_{Ai} and l_{Cu}, with 0 °C? (Linear expansivities of aluminium and copper are 2.3 x 10-5 °C^{-1} and 1.7 x 10-5 °C^{-1} respectively.)**



Expansion of Solid

It is a simple question. $l = l_0(1 + \alpha\theta) \rightarrow \frac{l-l_0}{l_0} = \alpha\theta$; When the fractional increment $\frac{l-l_0}{l_0}$ is plotted against you will get a straight line which goes across the origin. The gradient is equal to As the linear expansion of Al is greater than the value of copper (Cu), the gradient of the straigth line for Al should be greater. Therefore, the correct variation is (1).

Think in a simpler way. There cannot be a change in length if there is no temeprature change. Therefore, when $\theta = 0$ then the fractional increment or the increment in length should be zero. So, when $\theta = 0$, there cannot be an intercept in the straight lines. (4) and (5) can be removed. Even same straight line cannot be obtained for Al and Cu. From that (3) is removed. As the α of Al is greater than α of Cu, the straight line with a greater gradient should go to Al.

20. During the recent hot season, the night time temperature of a certain room with closed windows in a house made of bricks was observed to be 35 °C. A person opened the windows of the room for a few minutes at night and allowed the room to be filled with cooler air at 27 °C which was present outside the house. Once the windows were closed again, he observed that the temperature of the room had returned almost to 35 °C in a quick time. Which of the following reasons he had proposed to explain the observed effect is most unlikely to be accepted?

(1) Rapid movement of air molecules inside the room

(2) Collision of air molecules with the walls

(3) Low specific heat capacity of air

(4) Low thermal conductivity of air

(5) High specific heat capacity of brick walls

It is a beautiful question. If the air of 27°C needs to be 35°C air, then the air molecules should get the heat by the colliding with the walls. There is no other source which can give the heat. As there is higher specific heat capacity in the bricks, the walls get heated and store some amount of heat during the day time. When the gas molecules are hit with bricks, the heat gets transferred to the gas molecules. If we consider that the wall is in 35°C, then the gas molecules in 27°C will hit on the wall and get the heat.

If the motion of the gas molecules in the room happens very quickly, then the heat transfer from the wall also happen very quickly. When the number of collisions with wall increases, the heat transfer happens quickly. As the specific heat capacity of the air is less, the air gets hotter from a small a heat amount. If a certain heat is given to a material with a less specific heat capacity, then the temperature increases to a considerable level. If a certain heat was supplied to a material which has a great specific heat capacity, then the temperature will not rise that much. Likewise, when it gets cooled, it does not gets cooled quickly.

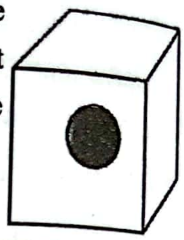
According to the clarification of the observations, the unaccepted reason is the less heat conductivity of the air. Heat is transferred by conduction from one molecule to the other molecule not by moving of molecules from one place to another. Therefore, if the air molecules are heated due to heat conductivity, it should happen first from the molecules near the wall towards to the molecules in the room. As the gas molecules are in random motion, the heat transfer from the conduction is not efficient. Therefore, the lowest bid to explain the observation should be kept to the less heat conductivity of the air.

Even from the five answers, the unsuitable answer is the heat conductivity. The other four answers are inter-connected. We normally do not say that a gas gets warmer or cooler due to heat conduction. The molecules of a solid material are placed closer to each other. As these molecules are doing vibrations in a certain limit, the energy can be transferred from one molecule to another molecule (conduction). But as the distance between the gas molecules are greater, the energy transfer from molecule to molecule does not occur regularly.

The mean kinetic energy of air molecules in 35°C is higher than the value in 27°C. The air molecules with higher kinetic energy have a higher possibility to diffuse to the outside.

The window is being opened for few minutes has been mentioned in the question because if it is opened for a longer time, due to the diffusion of air molecules the room temperature will acquire the outside atmosphere temperature of 27°C. But the question does not mention about a heat exchange like this way. However, if the inserted air molecules of 27°C to the room is needed to reach 35°C, then the heat should be gained from the inner walls. That means by only colliding with the inner walls. Otherwise, there is no other way for the air molecules to get the heat. There is no other heat source in the room to provide heat to the air molecules of 27°C that came into the room. The air molecules collide with the walls and absorb the heat of the walls. This should be an inside game as the windows are also being closed here.

21. A cube of ice of mass 1 kg at °C has a small metal sphere trapped inside as shown in the figure, It was found that this ice cube requires 300 kJ of heat energy to completely melt and form water at 0 °C. Specific latent heat of fusion of ice is 330 kj/kg, The mass of the metal sphere in grams is approximately



(1) 30 (2) 33 (3) 91
(4) 110 (5) 333

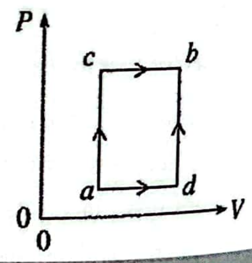
Calorimetry 04

You need a calculation. The initial temperature of the cube is 0°C. By supplying heat, 0°C ice is converted to 0°C water. What is indicated here is that the temperature of the metal sphere is remaining at 0°C. Therefore, the supplied heat is totally used to covert ice into water. The metal sphere does not absorb any heat. The mass of only ice = 300/330 kg (Q =m L); Therefore, the mass of the metal sphere = 1- 30/33 =3/33 = 1/11 kg. When it is converted to grams, (1/11) X 1000 g $\cong 91\ g$.

The answer is being asked to a near value. That indicates that it does not simplify accurately. If there is 10 instead of 11, the answer is 100 g. Therefore, the answer should be less than 100. There is only one answer which is less than 100 and near to that value. It is 91. So, you can get the answer without dividing 1000 by 11.

The trick of this question is to know that the metal sphere does not absorb heat. Even the specific heat capacity of the metal is not given. 1kg should be taken as the total mass of the metal and ice. However, according to the given data, the mass of only ice is less than 1 kg.

22. An ideal gas is taken from state *a* to state *b* through two paths *acb* and *adb* as shown in the $P-V$ diagram. When going through path *acb*, 100 J of heat is absorbed and 50 J of work is done by the gas. If the work done by the gas, when taking the path *adb* is 10 J, the amount of heat absorbed by the gas during the path *adb* is



(1) 40 J (2) 50 J (3) −50 J
(4) 60 J (5) −60 J

Thermodynamics 04

Such questions can be seen in the past papers a lot. As it is an ideal gas, if the initial and final states are same even it goes from any path, then the internal energy difference of the gas will be same. Therefore, (ΔQ-ΔW) is constant for both paths.

100- 50 = ΔQ – 10; ΔQ = 60 J.

This is all that you need to write as rough work. In both instances, the work is done from the gas. Therefore, ΔW is positive. As the heat is absorbed, ΔQ is also positive. The answer for ΔQ will be a positive value. Therefore, even in that path, the heat is absorbed.

23 If the ratio, $\frac{\text{mass of the planet}}{\text{radius of the planet}}$ for planet A is four times that of planet B, then the ratio $\frac{\text{Escape. velocity at the surface of Planet A}}{\text{Escape velocity at the surface of Planet B}}$ is

(1) $\sqrt{2}$ (2) 2 (3) 4 (4) 8 (5) 12

Gravitation Force Fields **05**

There are questions of escape velocity in the previous papers. $v = \sqrt{\frac{2GM}{R}}$. This expression is being asked in the 21st question of paper 2013. $v \propto \sqrt{M/R}$

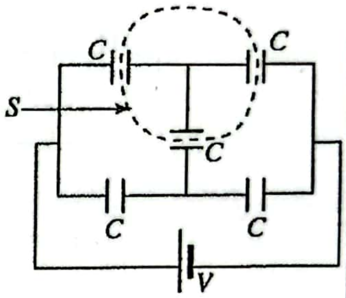
Therefore, $\frac{v_A}{v_B} = \sqrt{\frac{M_A/R_A}{M_B/R_B}} = \sqrt{4} = 2;$

It has been given that $\frac{M_A}{R_A} = 4\ \frac{M_B}{R_B}$. 4 is given as the square root of 4 is 2.

If the mass is M and radius is R, then the minimum escape velocity of v from a planet for a particle is given by

1) $v = \sqrt{\frac{2GM}{R}}$ 2) $v = 2\sqrt{\frac{GM}{R}}$ 3) $v = 4\sqrt{\frac{GM}{R}}$ 4) $v = GM/R$ 5) $v = 2GM/R$

24. A network consisting of five identical parallel plate capacitors of capacitance C each, is connected to a cell of voltage V as shown in the figure. Assume that the capacitor plates are in free space. The net electric flux through the enclosed surface S is



(1) $\frac{CV}{2\epsilon_0}$ (2) $\frac{3CV}{5\epsilon_0}$ (3) $\frac{CV}{\epsilon_0}$

(4) $\frac{3CV}{\epsilon_0}$ (5) 0

Capacitance and Capacitors **06**

You know that the charges are not stored in the capacitor which is in the middle. Look at the 14th question paper 1990. Even in a similar resistor arrangement, you know that there will be no current flow across the middle resistor (Wheatstone circuit arrangement). The upper two capacitors are in series with each other. If you consider the charges in the plates, then the net charge in the closed S surface is zero = -q + q + 0 = 0. Therefore, the net electric flux is zero across S surface. Does q/ε_0 or q is the electric flux? According to the SI unit of electric flux, the unit is C (Coulomb). According to the official web page for SI units http:// physics.nist.gov/cuu, the unit of the electric flux density is Cm^{-2} (the electric flux across a unit area). The section which has this information has been shown here. But in many international books in advanced level have mentioned q/ε_0 as the electric flux. It is true. The examiners have decided to take q/ε_0 as the electric flux from this onwards. In the new syllabus of Physics from 2017 also has decided to interpret electric flux as q/ε_0 not as q. Even though my opinion is different, all the others opinion is q/ε_0 not q. It is better to come into a common standard for a certain way. We must respect the decision of the majority.

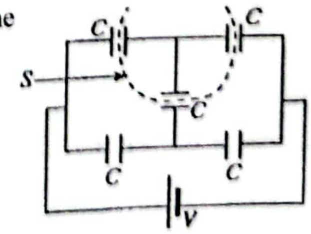
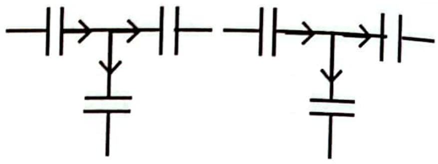
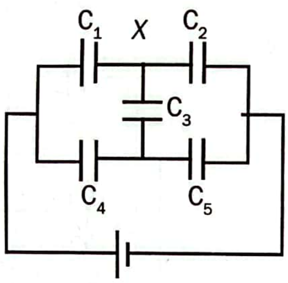


Table 4. Examples of SI derived units whose names and symbols include SI derived units with special names and symbols

Derived quantity	SI derived unit: Name	SI derived unit: Symbol
dynamic viscosity	pascal second	Pa s
moment of force	newton meter	N·m
surface tension	newton per meter	N/m
angular velocity	radian per second	rad/s
angular acceleration	radian per second squared	rad/s^2
heat flux density, irradiance	watt per square meter	W/m^2
heat capacity, entropy	joule per kelvin	J/K
specific heat capacity, specific entropy	joule per kilogram kelvin	J/(kg·K)
specific energy	joule per kilogram	J/kg
thermal conductivity	watt per meter kelvin	W/(m·K)
energy density	joule per cubic meter	J/m^3
electric field strength	volt per meter	V/m
electric charge density	coulomb per cubic meter	C/m^3
electric flux density	coulomb per square meter	C/m^2
permittivity	farad per meter	F/m
permeability	henry per meter	H/m
molar energy	joule per mole	J/mol
molar entropy, molar heat capacity	joule per mole kelvin	J/(mol·K)
exposure (x and Ƴ rays)	coulomb per kilogram	C/kg
absorbed dose rate	gray per second	Gy/s
radiant intensity	watt per steradian	W/sr
radiance	watt per square meter steradian	W/(m^2·sr)
catalytic (activity) concentration	katal per cubic meter	kat/m^3

Even though the capacitors in this question are not identical in capacity, the answer will not be changed. A question with unequal capacitors has been asked in the 59th question of paper 1984. As a sentence of this question, it was given that the algebraic sum of the charges of capacitor plates that are connected to the point X is zero. This sentence is correct. This can be proved like this way. When the capacitors are being charged by connecting to the cell, there will be a current flow in the branches till the capacitors are totally charged and get into the continuous state. The rate of charge flow with time means a current flow.





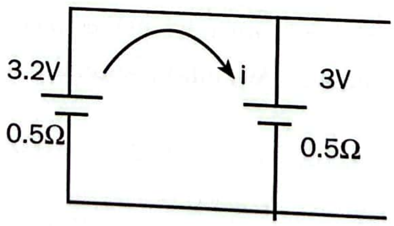
At a certain instance, if the currents that flow in the branches are i_1, i_2 and i_3, then we know that $i_1 = i_2 + i_3$. That means $i_2 + i_3 - i_1 = 0$. According to this $q_2 + q_3 - q_1 = 0$. The algebraic sum of the charges is zero. Therefore, to get the net electric flux across the surface S as zero, it is not definitely needed to have same capacity values for the capacitors. This is true for any occasion. But when the capacity values of the capacitors are equal, then the question is simple.

25. Two cells having e.m.f.s of 3 V and 3.2 V and equal internal resistances of 0.5 Q are connected in parallel as shown in figure. Power dissipation by the cell combination is

(1) 0.01 W (2) 0.02 W (3) 0.03 W
(4) 0.04 W (5) 0.05 W

Korchoff's Law - Combinations of Cells 08

It is simple. If needed, you can do it from your memory. If the flowing current is i, then 3.2 - 3= *i* (0.5+0.5);i=0.2 A. then The power produced from the combined system = $(0.2)^2$ X 1 = 0.04 W. I told you that you can do it from your memory due to this reason. The net e. m. f around the network is 0.2 V. The total resistance is 1 Ω. Even the current flows to any direction, there will be a power consumption across a resistor when there is a current flow. The square of 0.2 is 0.04. You might take the square of 0.2 as 0.4. But there is no answer for 0.4.



26. Nine identical wires made of a certain metal, each of diameter d and length L, are connected in parallel to form a single resistor. The resistance of this resistor is equal to the resistance of a single wire of length L and diameter D made of the same metal if D is equal to

(1) $\frac{d}{3}$ (2) $3d$ (3) $6d$ (4) $9d$ (5) $18d$

Ohm's Law Combination of Resistance **08**

You can get the answer very easily from my proportionality method. The metal has not been changed. The length (l) has not been changed. Therefore, the resistance of the wire R with a diameter $R \propto \frac{1}{d^2}$ $(R = \frac{\rho l}{A})$. If such nine wires are in parallel, then the net resistance $\propto \frac{1}{9d^2}$. When the resistors are parallel, the equivalent resistance gets reduced. It cannot get increased. When there are more alternatives, the resistance is reduced. When there is a trouble, the stress can be reduced if there are many alternatives.

The resistance of the wire with a diameter $D \propto \frac{1}{D^2}; \frac{1}{D^2} = \frac{1}{9d^2}; d = 3d$.9 is given as 3 is its square root. If you think in another way, then nine wires are being parallel means that the cross-sectional area of a single wire has been increased by 9 times. Therefore, to get that cross- sectional area of 9 times, there should be 3 diameters ($3^2 = 9$).

27. A structure consisting of straight wire sections of AO, OB, CO and OD of equal lengths arranged so that AOC = BOD , carry currents I along the directions shown. When this structure is placed perpendicular to a magnetic field as shown in the figure, due to magnetic field it will experience

(1) a resultant force along the plane of the paper in the upward direction.

(2) a resultant force along the plane of the paper in the downward direction.

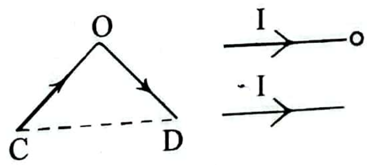
(3) a resultant force along the plane of the paper to the right.

(4) a resultant force along the plane of the paper to the left.

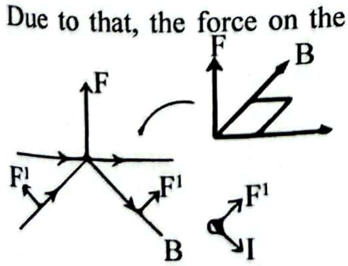
(5) no resultant force.

Forces on a Moving Charge in a Magnetic Field **07**

Look at the 31st question of paper 2013 and 33rd question of paper 2011 (old syllabus). The force on a current carrying wire with any shape at a uniform magnetic field is equal to the force which is applied if the two ends of the wires are connected by a straight wire. Therefore, the wire part of OCD is equal to the wire part of CD which is shown in a dashed line. Now the system has two horizontal wire parts which carry the current to the right side.

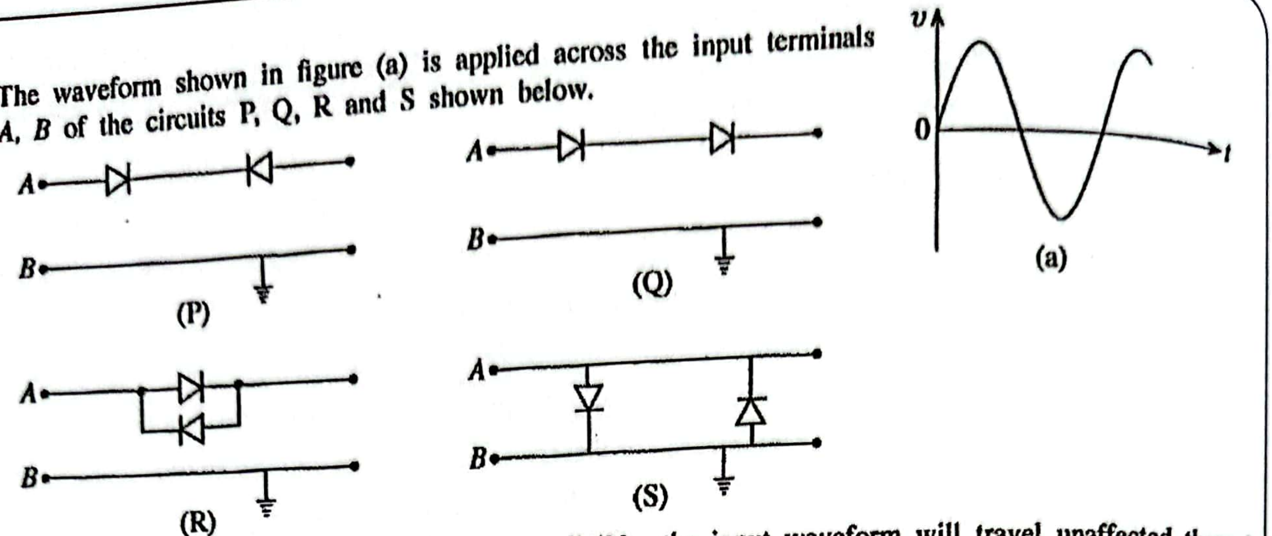


Now keep the right thumb perpendicular to the other fingers and rotate these fingers from I to B. Then the thumb is pointing upwards. The answer is (1). The question asks about the force on the current carrying wires due to the external magnetic field. Also, there is a magnetic field due to the current flow of the wires. Due to that, the force on the wires is attracted to each other. This system is equivalent to two parallel wires which carry current to the same direction. Even you can solve the question using this fact. The acting forces on the system can be drawn according to the way the system is drawn. Look at the figure. When F' is resolved into horizontal and vertical directions, the horizontal components are cancelled off whereas the components that are acting vertically upwards are left out.



S R D ROSA I SciTech.LK

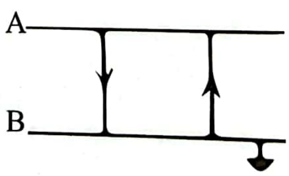
28. The waveform shown in figure (a) is applied across the input terminals A, B of the circuits P, Q, R and S shown below.



If the potential drops across the diodes are negligible, the input waveform will travel unaffected through

(1) the circuit P only.
(2) the circuit Q only.
(3) the circuit R only.
(4) the circuit S only.
(5) the circuits R and S only.

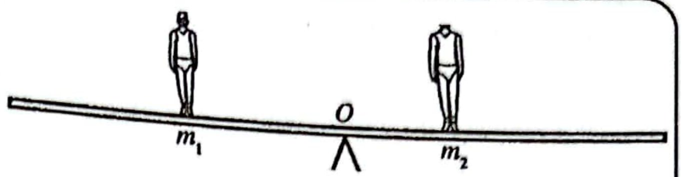
The road has been opened and it has been closed again in P. In the positive part of the wave model, the first diode is forward biased. Therefore, the positive part of the model goes across that diode. But the second diode is reverse biased for the positive part of the wave model. Likewise, the first diode is reverse biased for the negative part. Actually, nothing goes across the circuit P. the road has been opened/closed and then closed/opened. From Q the positive part of the wave model is only passed. The negative part is cut off. At a glance (P) and (Q) can be removed and also (S) can be removed. The diode ends in (S) are being earthed. Therefore, you do not have to look at (S). Even the positive part of the wave model tends to go across the first diode of (S), the output is earthed. In another way, the circuit is being short circuited. As the potential drop across the diodes is not considered, the diodes can be treated as equivalents of two wires which flow current to one side if needed. Then (S) circuit is like a circuit where the live wire is connected to the earth wire which is short circuited. (R) is correct. The two diodes are connected to both sides in parallel. The other diodes are on the same road. The positive part of the wave goes across the upper diode whereas the negative part goes across the lower diode. As the positive part is allowed from the top and the negative part is allowed to flow from the bottom, the wave model flows without an effect.



One gate is open when needed. Then the other gate is closed by that time. At the other time the previously opened gate is closed whereas the previously closed gate is opened now. This work cannot be done by connecting the diodes in series. If we need to insert two groups separately to a hall according to our wish, then we need two separate doors. At the polling centres there are separate queues for males and females. Both enter separately in two doors. If the doors are placed in series, then this work cannot be done. Finally, the ballot is put into one box. The correct answer is (3). The earthed connection is shown in these circuits to show that the voltages are measured relative to the earthed (0) connection. The earthed connection is acting as a reference point/ way.

If such a wave model is used in cathode ray oscilloscope (CRO) to observe, then the earthed connection is the metal chasis/frame of the CRO.

29. Two children of masses m_1 and m_2 are standing in equilibrium as shown in figure, on a uniform rod which is balanced at its centre of gravity O. Then they start moving simultaneously on the rod at constant speeds v_1 and v_2 respectively while maintaining the horizontal equilibrium of the rod. Consider the following statements made about the motion of the two children. For the equilibrium to be maintained at any time t,



(A) they should always move in opposite directions.
(B) they should move keeping their total linear momentum always equal to zero.
(C) they should move so that the moment produced by one child about O is always equal and opposite to the moment produced by the other child about O.

Of the above statements,

(1) only A is true.
(2) only B is true.
(3) only A and B are true.
(4) only B and C are true.
(5) all A, B and C are true.

Newton's Law and Momentum **02**

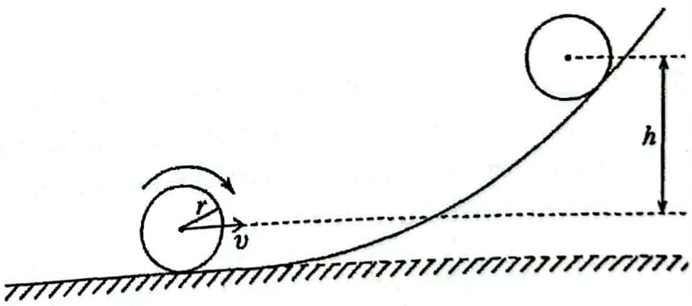
All you need is to keep the mass centre/ centre of gravity of the system at the point O always. That means always it should be $m_1r_1 = m_2r_2$. If r_1 is increased, then r_2 also should be increased. Likewise, if r_1 is decreased, then r_2 should be decreased. Therefore, two children definitely should move into opposite directions. That means $\leftarrow$ *and* $\rightarrow$ *or* $\rightarrow$ *and* $\leftarrow$. It cannot be $\leftarrow$ *and* $\leftarrow$ *or* $\rightarrow$ *and* $\rightarrow$ Therefore, statement (A) is true. It can be decided that this statement is true from general knowledge too. If it does not move in the opposite directions, then the balance will be upset. (B) can be obtained from $m_1r_1 = m_2r_2$. If this expression is divided by time (t), then $m_1r_1/t = m_2r_2/t$; $m_1v_1 = m_2v_2$; $m_1v_1 - m_2v_2 = 0$. (B) is also true. Also, (C) can be obtained from $m_1r_1 = m_2r_2$. $m_1gr_1 = m_2gr_2$. From general knowledge you can see that (C) is true.

The direction of the moment is outside the paper. $\odot$

The direction of the moment is into the paper. $\otimes$

The thumb of the right hand is kept perpendicular to the other fingers and the fingers are rotated from vector r to F. Then the direction that the thumb points to the direction of the moment vector. The clockwise and anticlockwise directions are not the vector directions of the moment. These are the directions that we made for our convenience.

30. A uniform disc of mass m and radius r rolls without slipping, initially along a horizontal surface, and subsequently starts to climb up a ramp as shown in the figure. The disc has a linear velocity v on the horizontal surface. The moment of inertia of the disc about the axis through its centre and normal to the plane of the disc is $\frac{mr^2}{2}$. What is the maximum height h to which the centre of mass of the disc climb?



(1) $\frac{v^2}{2g}$ (2) $\frac{3v^2}{2g}$ (3) $\frac{3v^2}{4g}$ (4) $\frac{v^2}{g}$ (5) $\frac{2v^2}{g}$

Work Power and Energy **02**

We need to write an equation (conservation of energy). It is a familiar situation. The disk has a linear and a rotational kinetic energy. Rolling is happened in the disk. A rolling can be considered as the total of pure linear motion and a pure rotational motion. This fact has been reviewed in earlier papers.

Initially, the total of linear kinetic energy and the rotational kinetic energy should be equal to mgh. At the maximum height, there is no linear or rotational motion. The total kinetic energy is transformed into gravitational potential energy. When writing equations, try to write them in one row. The velocity of the mass centre of the disk should be v.

$\frac{1}{2} mv^2 + \frac{1}{2} (mr^2/2) \times v^2/r^2 = mgh$; $h = \frac{3}{4} v^2/g$; $[1/2\ I\omega^2 = \frac{1}{2} (mr^2/2) \times v^2/r^2]$; When ½ and ¼ are added it is ¾.

If the disk is sliding, then v cannot be written as $v = r\omega$. If the surface is smooth, then the disk will slide. Friction is needed if the disk is needed to be rolled. There will be a created torque at the centre of the disk due to the frictional force. But as it is not sliding, there will not be any work done from the frictional force.

31. A glass of fresh orange solution of volume 500 cm^3 contains a few orange seeds at its bottom. It was observed that the seeds just began to float at the bottom when 10 grams of sugar was dissolved in the solution. Assume that the addition of sugar does not alter the volume of the solution. If the density of the orange solution before adding sugar was 1000 kg m^{-3}, the density of orange seeds (in kg m^{-3}) is approximately equal to

(1) 1020 (2) 1040 (3) 1060 (4) 1080 (5) 1100

Hydrostatics **02**

The orange seeds are starting float means that the density of the orange seeds have become equal with the density of the liquid. The work is done when we find the density of the solution after sugar is applied. As there are g, kg, cm^3 and m^3, you can get jumbled. If the density of the water is converted to gcm^{-3}, then the calculation is easy. 1 gcm^{-3} is 10^3 kgm^{-3}.

$10^3\ kg/m^3 = 10^3 \times 10^3\ g/cm^3 = (10^3 \times 10^3)/10^6\ g/cm^3 = 1\ gcm^{-3}$

When we were doing A/Ls, we remember the density of water as 1 gcm^{-3}. Those days, the questions were done in g and cm^3.

If 1g is 1cm^3, then 500 cm^3 has a mass of 500g.

As 10g is put to the water, now the mass of the solution is (500+ 10) 510 g. The volume of the water does not change as sugar is inserted. So, the volume of the solution is 500 cm^3.

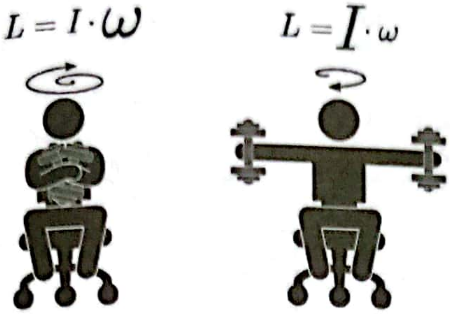
Therefore, the density of the solution = 510/500 gcm^{-3} = 1. 02 gcm^{-3}. That means 1020 kgm^{-3}. The correct answer is (1). To convert gcm^{-3} to kgm^{-3}, you need to multiply by 10^3. But if g is needed to make kg, then you need to divide from 10^3. As there is 10^6 in m^3, m^3 also should be multiplied by 10^6. Therefore, $10^6/10^3 = 10^3$. Even the division of 51 from 5 should be seen as 10.2 for the answer. So, other answers instead of 1020 are irrelevant.

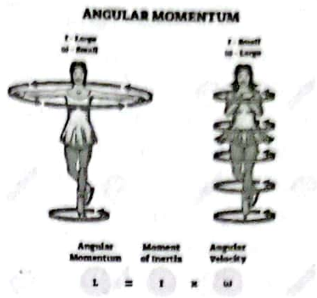
32. A boy, sitting on a smooth turntable with a weight in his each extended hand, is rotating with an angular velocity w0. When he bends his hands towards his body, the angular velocity becomes cop If /0 and /1 are the moments of inertia of rotating systems when the hands are extended, and bent towards his body respectively, then

(1) $\omega_0 > \omega_1$, $I_0 > I_1$ and $\omega_0 I_0 > \omega_1 I_1$ **(2) $\omega_0 < \omega_1$, $I_0 > I_1$ and $\omega_0 I_0 < \omega_1 I_1$**
(3) $\omega_0 < \omega_1$, $I_0 > I_1$ and $\omega_0 I_0 = \omega_1 I_1$ **(4) $\omega_0 > \omega_1$, $I_0 < I_1$ and $\omega_0 I_0 = \omega_1 I_1$**
(5) $\omega_0 = \omega_1$, $I_0 = I_1$ and $\omega_0 I_0 = \omega_1 I_1$

Work Power and Energy **02**

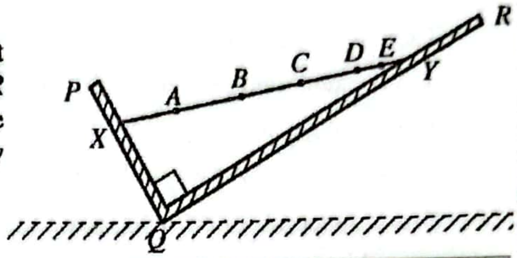
It is very simple. This has been checked in several situations. Look at the 49th question of paper 2012 (old). The angular momentum has to be conserved. That means $I_0\omega_0 = I_1\omega_1$. Choices of (1) and (2) are removed. When the hands are spread, the moment of inertia gets increased. Therefore, $I_0 > I_1$. If $I_0 > I_1$, then $\omega_0 < \omega_1$. The correct answer is (3). When the hands are moved towards the body, we know that the angular velocity increases. The videos of people who are doing ice skating on television come into the mind effortlessly. The child is holding two loads from each hand.





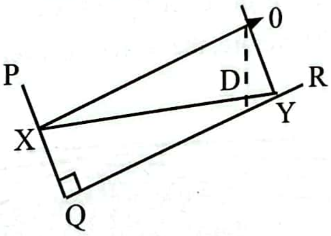
33. A rod XY rests between two smooth boards PQ and QR kept inclined to the horizontal as shown in the figure. Angle PQR is 90° and the surfaces of the boards are normal to the plane of the paper. The centre of gravity of the rod is most likely to be situated at the point

(1) A (2) B (3) C
(4) D (5) E

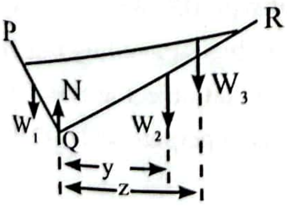


You need to consider that the plates are kept at rest without motion. The question mentions that the plates are kept according to the figure. Therefore, do not ask how these plates are kept in this way. The location of the centre of gravity of the rod is being asked. If we consider the forces acting upon the rod, then there are three forces. The weight of the rod, the perpendicular reactions on the rod by the plates at the points of X and Y. The plates are smooth. Therefore, you need to draw perpendicular lines from X and Y points to the plate and decide the place where they intersect. The action line of the weight of the rod should go across that point (equilibrium of three forces).

As $P\hat{Q}R = 90^0$, the perpendicular reactions should be parallel to the opposite plates. That means XOYQ is a rectangle. As QY> XQ, the O corner of the rectangle, cannot be straight up to the points of A, B and C. Point E is also very closer to Y end. Therefore, the correct point is D.



If we consider the plates and the rod as a system and if the system has to be in equilibrium, then the vertical line which goes across the centre of gravity should across point Q. The plates are not fixed at Q point has been used as a hypothesis. If so, then plate PQ should be heavier than plate QR. As the length of plate QR is greater than PQ, both plates should not be made from the same material to satisfy this fact. QR should be made from a light material.



The figure has shown the forces of the system. W_1 – the weight of PQ plate W_2 – the weight of PQ plate W_3 – the weight of the rod.

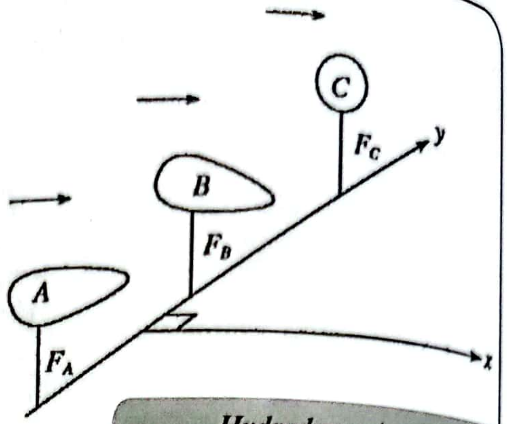
If $W_1 x = W_2 y + W_3 z$, then the system can be kept nearly in equilibrium. But it is not a stable equilibrium. It will roll once it is moved slightly.

34. Two objects A and B of the shapes shown in the figure, and a spherical object C, all having identical masses, are mounted rigidly on a horizontal surface along the y-axis by three thin rods as shown in the figure. Both x and y axes are located on the horizontal surface.

A stream of air flows through the objects parallel to the surface and along x-direction. (Assume that the air flow causes no turbulence around the objects.) The magnitudes of the forces F_A, F_B and F_C exerted by the objects and the sphere on the mounted rods, when written in the ascending order, it will be

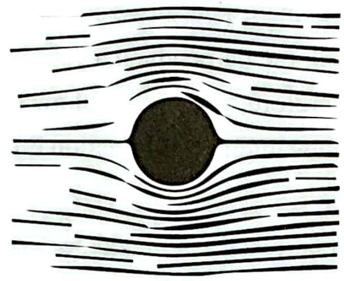
(1) F_B, F_A, F_C (2) F_B, F_C, F_A (3) F_C, F_A, F_B

(4) F_A, F_C, F_B (5) F_C, F_B, F_A

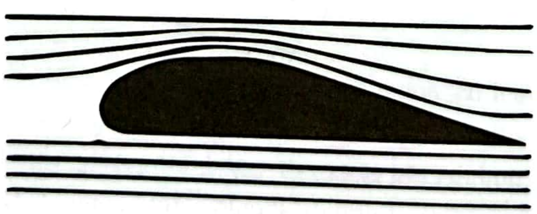


There is an answer if you do not think far. When you see the shape of A and B objects, it is a surprise that if you cannot understand that this should be a question related to Bernoulli theorem. As C takes a spherical shape, the gas flow lines flow uniformly from the top and the bottom. Look at the figure. Therefore, there is no net force to upwards or downwards on C from the gas flow lines.

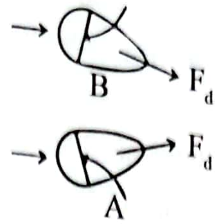
B is a shape of an air wing as we all know (wing of an aeroplane). When you consider the streamlines around it, everyone knows that there is a created net upward force (lift). The object A is like the upside-down version of the object B. Therefore, when streamlines are flown across A, the pressure near the bottom surface is lesser than the flat surface. The speed of the streamlines across the bottom surface is greater compared to the streamlines of upper surface. Therefore, there is net downward force on A.



B tends to lift upwards and A tries to push downwards. There is no change in C. Therefore, $F_B < F_C < F_A$. As it tries to go upwards, the least force is there for F_B. As A tries to go downwards, the highest force is there for F_A. C is doing a neutral duty. Therefore, the correct answer is (2).



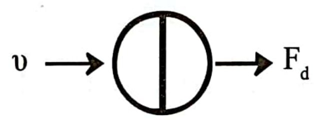
Why this was given ALL? When objects are moved across the air flows, there is a drag force on the objects. Can you remember the first essay question (5th question) of paper 2015? The drag force is proportional to the flow and the object's frontal area ($F_d = \frac{1}{2} c_d d A v^2$). Here the air flow is hit with the objects and forces are being created. That force even can be obtained from the above relation. An object travelling in v speed at still air compared with the flow of air by v speed to the opposite direction on an object at rest is equal according to Physics principles. Here d = the density of air and A = the frontal area perpendicular to the air flow.

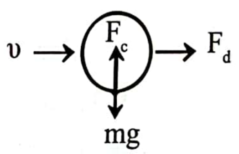


According to the way that A and B objects are drawn, their frontal areas are not perpendicular to the flow lines. Therefore, the force from the gas actually does not act horizontally. To find the force, the relevant frontal area should be multiplied by the square of the perpendicular component of the gas velocity to that area. As the applied force F_d is acting in an inclined way, their components are acting on vertically downwards (at B) and vertically upwards (at A). Therefore, these component parts are affecting the forces of F_A and F_B. To find the force on the spike of B, the vertically downward force of F_d should be added. Therefore, we cannot accurately decide whether F_B is the least force. Due to Bernoulli effect B tries to move upwards. But from the force component from the gas, B gets a downward force. If the force component from the gas is acted vertically upwards, then the work is ok. If so, then due to the lifting force and the force component from gas F_B will be the smallest force. Likewise, you can argue for A.

Due to Bernoulli effect, there is a vertical downwards force on A and the force component from the gas is vertically upwards. The directions are opposite to each other. Therefore, you cannot say which force is greater in magnitude. If the force component from the gas is acted downwards, then from both of the forces F_A will be the largest force.

Think if the frontal area is perpendicular to the flow. Then there will not be an issue.

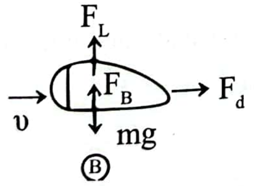


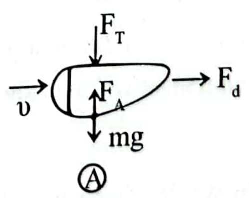


For C, $F_C = mg$. For C, however the force from the gas is horizontal.

For B, F_L = lifting force; mg = the weight of the object; F_B = the force on the object by the spike. The force on the spike by the object is equal and opposite to this force. Now $F_B = mg - F_L$ (there is no vertical component from F_d)

For A, $F_A = mg + F_T$

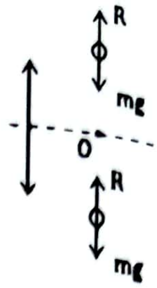




35. A mass is resting on a horizontal surface which moves up and down performing simple harmonic motion with amplitude A as shown in figure. The maximum frequency with which the surface can move while keeping the mass always in contact with the surface is

(1) $2\pi\sqrt{\frac{g}{A}}$ (2) $\sqrt{\frac{g}{A}}$ (3) $\frac{1}{2}\sqrt{\frac{g}{A}}$ (4) $\frac{1}{2\pi}\sqrt{\frac{g}{A}}$ (5) $\frac{1}{\pi}\sqrt{\frac{g}{A}}$

When I see this question, I remember my past. The first structured question of paper 1974, which is the year I did my A/L was built based on this question. If the object was not thrown at the highest end of the amplitude, then it will be not thrown at any other place. Both go together. We will consider the forces on the object.



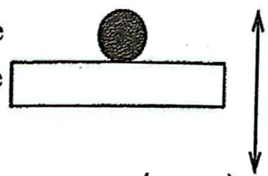
There is a downward acceleration of $\omega^2 A$ (towards the centre) at the highest end of the amplitude. The acceleration of an object which under goes simple harmonic motion a = - $\omega^2 x$ where x is the displacement. Now when we apply F = ma downwards, then mg - R = m $\omega^2 A$;

The maximum value of ω that is just needed to get R zero will be $g = \omega^2 A = (2\pi f)^2 A$; $f = \frac{1}{2\pi}\sqrt{\frac{g}{A}}$

This is equivalent to the removal of paper pieces when the wires of the sound measuring instrument are having resonance. Look at the review of the 3rd structured question in paper 2007. When we apply F = ma upwards to the lowest point of the amplitude, then R - mg = m $\omega^2 A$

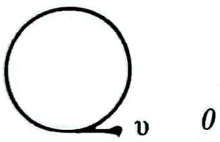
R cannot be zero at any instance. When you think this problem in a simpler way, the downward acceleration should be g if the reaction is needed to be zero. After standing in a lift, if the lift is falling at a downward acceleration of g, then the reaction of the feet and the bottom part of the lift is zero. Therefore, a = g = $\omega^2 A$. When it is converting to f, there should be a multiple of $1/2\pi$. So, the correct answer is (4).

36. whistle emitting a sound of frequency / moves along the circumference of a circle of radius r at a constant angular velocity m. I› is the velocity of sound in air. The highest frequency of sound heard by a listener, who is at rest outside the circle is



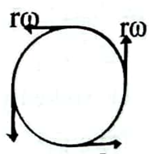
(1) $f\left(\frac{v}{v-r\omega}\right)$ (2) $f\left(\frac{v-r\omega}{v}\right)$ (3) $f\left(1-\frac{v}{r\omega}\right)$ (4) $f\left(\frac{v}{r\omega}\right)$ (5) $f\left(\frac{v}{v+r\omega}\right)$

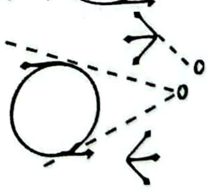
Doppler Effect 03



It is very simple. The highest frequency is heard by the observer O when the horn(source) is coming towards himself. You do not have to think about this fact too. According to $f' = f\frac{(v \pm v_o)}{(v \mp v_o)}$, here $v_o = 0$.

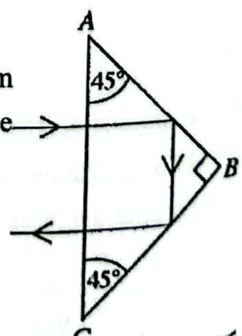
The source is coming closer to the observer. Therefore, $f' = f\frac{v}{(v - r\omega)}$. As the apparent frequency should be higher, the denominator should be (v - rω). It is there in only (1). In (2), (3) and (5), f'< f. If you stay at the centre of the circle, then the apparent frequency is f. The reason is that the speed component of the source is not directed towards the observer.





If the observer is at a different place, then the equation should be substituted with the velocity component which is towards or away from the observer.

37. A ray of light is incident perpendicular to ths surface AC of a right angled glass prism as shown in the figure. Minimum value of the refractive index of the material of the prism for which the ray will follow the path shown is



(1) 1.22 (2) 1.41 (3) 1.58
(4) 1.73 (5) 1.87

Doppler Effect 03

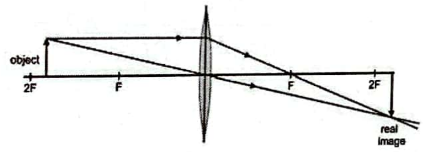
This is peanuts. Only 45^0 is given. Therefore, what else to do without taking 45^0 as the critical angle (c)? $n = \frac{1}{\sin 45^0} = \sqrt{2} = 1.41$. The ray that is subjected to total internal reflection has the incident angle of 45^0. Here the incident angle is taken as 45^0, which is the critical angle. As total internal reflection occurs at the incident angle of 45^0, it is ok if the critical angle is less than 45^0. But if the critical angle is less than 45^0, the relevant refractive index goes higher than 1.41. When c is reduced n is increased. Therefore, it is ok to have $\sqrt{2}$ as the minimum refractive index value. It is alright if it is greater than this value. If we take 1.5 as the refractive index of normal glass, then the relevant critical angle for this value is about 42^0. Then a ray with an incident angle of 45^0 however is subjected to total internal reflection. As 45^0 is given in the question, there is no other alternative than finding n for that value. $\sqrt{2}$ is the minimum value that n should have.

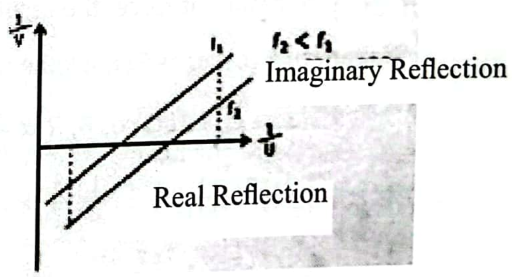
38. **When an object is placed on the principal axis of a thin convex lens of focal length f_1, it forms a real image at a distance V_1 with a linear magnification of m_1. When this lens is replaced by another thin convex lens of focal length f_2, ($f_2 < f_1$), being kept at the same position the new image distance V_2 and the magnification m_2 will satisfy the conditions,**

(1) $V_2 > V_1$ and $m_2 > m_1$
(2) $V_2 > V_1$ and $m_1 > m_2$
(3) $V_2 < V_1$ and $m_2 > m_1$
(4) $V_2 < V_1$ and $m_1 > m_2$
(5) $V_2 < V_1$ and $m_1 = m_2$

Refraction Through Lenses **03**

There is no need to write equations. The object distance has not been changed. When the focal length of the convex lens is reduced, the image distance also gets reduced for real images. That means $V_2 < V_1$. The magnification (m) is $|V/U|$. As U is unchanged, when V is decreased m is also decreased. Therefore, the correct answer is $V_2 < V_1$ and $m_1 > m_2$.





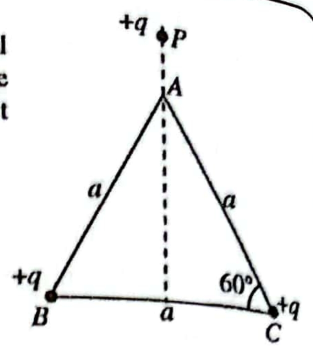
From the given ray diagram, you can quickly see that, when f is reduced V should be reduced as U is unchanged. A rough sketch is enough. The ray that goes across the optical centre does not change. Thank you to the teacher who had shown this to me.

Even from the graph of 1/U against 1/V this can be obtained.

$1/V - 1/U = 1/f \rightarrow 1/V = 1/U + 1/f$

As $f_2 < f_1$, then $1/f_2 > 1/f_1$. So, when the focal length is reduced, the numerical value of the intercept increases. The relevant graph goes down. Therefore, when $f_2 < f_1$, for a certain value of 1/U of real images 1/V value numerically increases. Therefore, numerically the value of V gets reduced. Even for unreal images when $f_2 <$ f_1, for a certain value of 1/U the relevant 1/V value gets reduced. That means the value of V gets increased. You can argue from the lens equation too. For real images, the lens formula is $-1/V - 1/U = -1/f$; $1/V = -1/U + 1/f$. So, when f is reduced, V should also be reduced as U is unchanged.

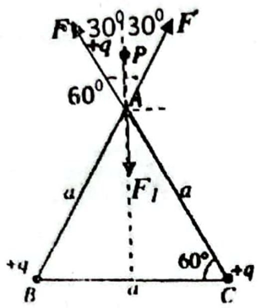
39. Two point charges of $+q$ each, are held at vertices B and C of an equilateral triangle ABC of side length a, and another point charge of $+q$ is held at the point P as shown in the figure. A zero resultant force will act on a positive unit charge placed at point A when the distance AP is equal to



(1) $\sqrt{2}a$ (2) $\frac{a}{2}$ (3) $\frac{a}{\sqrt{(\sqrt{3})}}$

(4) $\frac{a}{4}$ (5) a

This can be done in many methods.



Method 1

We will mark the forces on the unit positive charge on point A. There are three forces on it. The horizontal components of two forces of F are being cancelled off. Therefore, to get zero resultant force $2F \cos 30^0 = F_1$. There is no need to write full expressions for the forces. $1/4\pi\varepsilon_0$ is common.

Even q X 1 is common. So, $F \propto \frac{1}{distance^2}$. $F \propto \frac{1}{a^2}$; $F_1 \propto \frac{1}{(AP)^2}$; Therefore, 2 X $1/a^2$ X $\cos 30^0 = 1/AP^2$

$$\frac{2}{a^2}\frac{\sqrt{3}}{2} = \frac{1}{AP^2}; AP = \frac{a}{\sqrt{(\sqrt{3})}}$$

Method 2

To get zero resultant force, the resultant of two forces F should be equal and opposite to F_1. The resultant of two forces F should be numerically equal to F_1.

Apply $R^2 = P^2 + Q^2 + 2PQ\cos\theta$, $R \propto \frac{1}{AP^2}$; $P = Q \propto \frac{1}{a^2}$

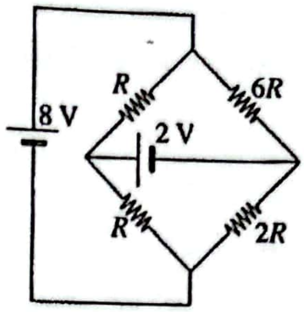
$$\frac{1}{(AP^2)^2} = \frac{1}{(a^2)^2} + \frac{1}{(a^2)^2} + 2.\frac{1}{a^2}.\frac{1}{a^2}.\cos 60^0; (AP^2)^2 = \frac{(a^2)^2}{3}; AP^2 = \frac{a^2}{\sqrt{3}}; AP = \frac{a}{\sqrt{(\sqrt{3})}}$$

Method 3

By using Lami's Theorem. Lami's theorem is not in Physics syllabus. So, biology students do not look at this method.

$F_1/\sin 60^0 = F/\sin 150^0 = F/\sin(90^0 + 60^0) = F/\cos 60^0 = 2/AP^2\sqrt{3} = 2/a^2 \rightarrow AP^2 = a^2/\sqrt{3}$

40. If you did not think of a shorter method, then the calculation will be lengthy. Do not apply Kirchhoff's equation by taking the current across 2V cell as i. The easiest method is to remove 2V cell from its place. Then the circuit will look like this way.



(1) a current of $\frac{3}{2R}$ passes through the 2 V cell.

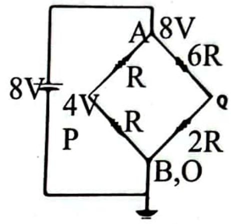
(2) a current of $\frac{6}{R}$ passes through the 2 V cell.

(3) a current of $\frac{10}{R}$ passes through the 2 V cell.

(4) a current of $\frac{3}{R}$ passes through the 2 V cell.

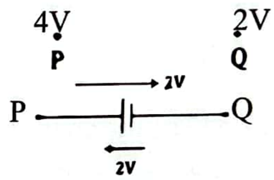
(5) a current does not pass through the 2 V cell.

40.If you did not think of a shorter method, then the calculation will be lengthy. Do not apply Kirchhoff's equation by taking the current across 2V cell as i. The easiest method is to remove 2V cell from its place. Then the circuit will look like this way.



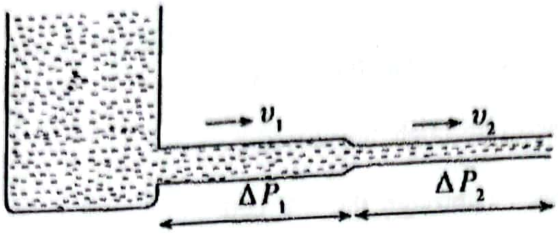
For convenience, earth the negative end. Then it is easy to find the potential values of relevant points. The potential of point B is zero. The potential of point A is 8 V. The total resistance of APB branch is 2 R. So, 8 V is divided equally among two Rs. Therefore, the potential of point P is 4 V (8-4 = 4-0). Now find the potential of point Q. The total resistance of AQB branch is 8R. The current that flows here is 8/8R = 1/R. The potential difference across QB = 1/R X 2R = 2V. Therefore, the potential of Q is 2 V. You do not have to find the current. It is just enough to divide 8 V to 3:1 (6 :2) ratio. Divide 8 by 4. Or else divide 8 by 8 and multiply by 2.

Now there is a potential drop from P to Q across PQ. Put 2V cell right now. What will happen? 2V and 2V are equal and opposite in direction. Therefore, there is no current across 2V cell. Both sides are equally matched. Many tend to think that there is 'no current flow' across the cell once they saw such a circuit. At this moment, that decision is against Physics.



If you try to find the currents in the branches as i_1, i_2, then it will consume lot of time. Therefore, use such methods to solve similar problems like these. Hope you can remember that in 10th question of paper 2015, the solving of the problem was easy when the wires are removed. Do not think this as a wrong thing. Think that the circuit was there without the cell of 2V. Then we can easily find the potentials of P and Q. Now think that if someone connected 2V cell across P and Q. As the previous potential difference of P and Q is equal and opposite to the e. m. f of the 2V cell, there is no change to the system. If 3V cell was put instead of 2V cell, then you can follow the above method. But now there is a current flow across 3V and the currents across the resistors change in the circuit. If so, this cannot be solved as a MCQ. Therefore, if such a question is given as a MCQ, then it should be given properly. R, R and 6R, 2R are given to make the potential difference across PQ as 2V.

41. **Two narrow tubes of equal lengths but different radi of cross-section are connected end to end, and water is allowed to flow through ii as shown in the figure.**



If v1 and v2 are the average velocities with which water flows through cross sections of the tubes and IP are the pressure differences built up across the tubrs as shown, then the ratio, $\frac{\Delta P_1}{\Delta P_2}$ is equal to

(1) $\left(\frac{v_1}{v_2}\right)^{\frac{1}{4}}$ (2) $\frac{v_1}{v_2}$ (3) $\left(\frac{v_1}{v_2}\right)^2$ (4) $\left(\frac{v_1}{v_2}\right)^3$ (5) $\left(\frac{v_1}{v_2}\right)^4$

Viscocity 10

It can be clearly seen that this is a question associated with viscosity. The length of the tubes is same with the same liquid. Therefore, cannot we write as $A_1v_1 = A_2v_2 = \Delta P_1 A_1^2 = \Delta P_2 A_2^2$? Av is the liquid volume that flows in a second. According to Poiseuille equation the flow rate (Av) = $\Delta P\pi r^4/8\eta l = \Delta P(\pi r^2)^2/8\eta l\pi$

Multiply and divide by π. The cross-sectional area is πr^2. As the multiple ηl is same the flow rate is proportional to ΔPA^2. So, $\Delta P_1/\Delta P_2 = A_2^2/A_1^2 = (v_1/v_2)^2$

The correct answer is (3). However, you should not expect a term with a power of 2. The square of r^2 is r^4 $[(r^2)^2]$.

42. **A student performed an experiment to verify the Boyle's Law using a constant mass md of an ideal gas at the room temperature of 27 °C and obtained the graph given in the figure. Here P is the pressure and V is the volume of the gas. He then removed a certain amount of gas from the volume *V* and repeated the expefiment at a temperature loll °C above the yoom temperature. tf the new s ph he obtained has the same gradient as the graph shown in the figure, the mass of the gas that he had removed is,**

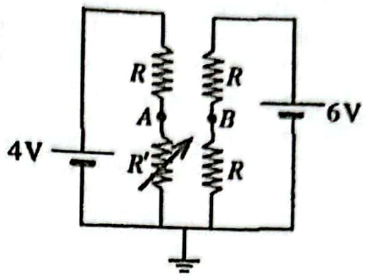
(1) $\frac{27}{100}m_0$ (2) $\frac{73}{100}m_0$ (3) $\frac{1}{4}m_0$ (4) $\frac{1}{2}m_0$ (5) $\frac{3}{4}m_0$

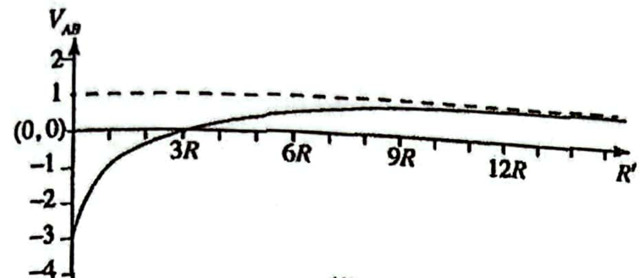
Expansion of Gases 04

It is simple. PV = nRT→ P =nRT. 1/V. Therefore, the gradient of the graph is nRT. As R is constant, if the same gradient is obtained for the two instances, then $n_1T_1 = n_2T_2$. As it is the same gas, the number of moles are proportional to the mass of the gas (n = m/W). So, $m_1400 = m_0300$ → $m_1 = 3m_0/4$. Therefore, the removee mass is ¼ m_o. There may be many students who chose ¾ mo. The removed mass is asked in the question. There is no secret that 0C temperatures should be converted to K. Always the room temperature is given as 27^0C to get 300 K (273 +27). When the temperature is increased by 1000C, then the new temperature is 400 in K. When the temperature is increased, if you need to the same gradient, then part of the gas should be removed. When T is increased n (m) should be decreased.

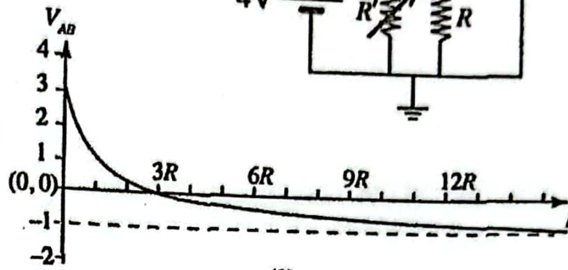
Look at this question which was given in 1981. This question has to be thought more than the question of 2016. The gases are different. The masses and the number of molecules are there. Can you argue that only A is correct? If they are in the same temperature, the moles of Y as well as the molecules should be greater than the values of X ($n_Y > n_X$). You cannot build the logic form the masses of the gases. We do not know anything about the molecular weight/atomic weight (molar weight) of them.

43. In the circuit shown, both cells have negligible internal resistances. R' is the value of a variable resistor. Variation of the voltage V_{AB} (= $V_A - V_B$) across the points A and B with R' is best represented by

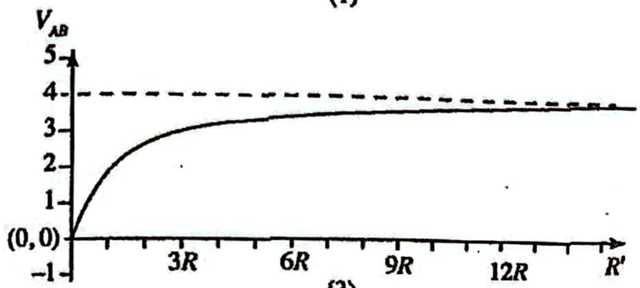




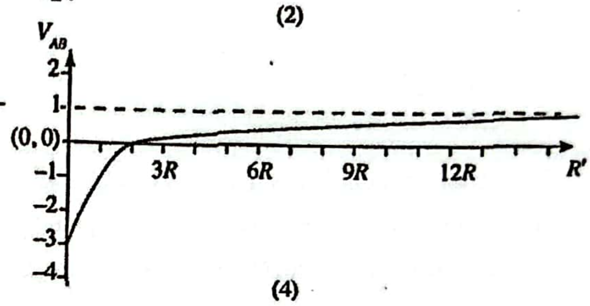
(1)



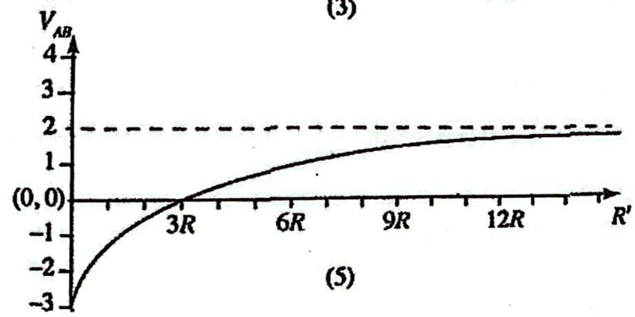
(2)



(3)



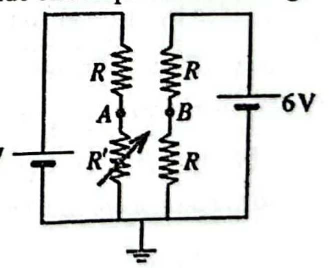
(4)



(5)

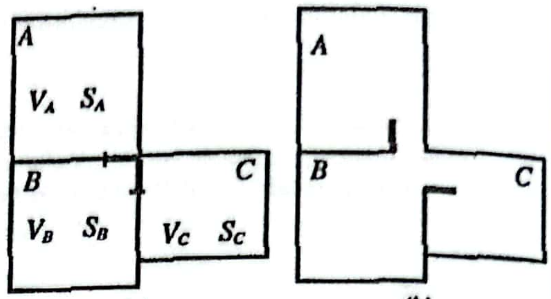
Potentiometer 08

Use the method that I have always taught. R' is being varied. V_B of the right side circuit part is not changed. Whatever happens to R', V_B is a constant. Is not it 3V? 6 V should be divided equally among R and R. Now make R' zero. Making R' zero means short circuiting the circuit. Putting a wire which has no resistance. Then A end gets earthed. That means $V_A = 0$. Then $V_A - V_B = 0 - 3 = -3$ V. So, the graph should start with -3V. Remove the graphs of (2) and (3). Now make R' as infinity. Making R' inifinite means breaking R'. the road is opened.



Then $V_A = 4$ V. There is no current flow in the left side of the circuit. The current is zero across R there. That means $V_A = 4$V. Now $V_A - V_B = 4 - 3 = 1$ V. Therefore, when Remove (5). Only (1) and (4) will be left. Think about the instance where $V_{AB} = 0$. When R' = 3R, what value will V_A take? When R' = 3R then $V_A = 3$ V. 4V is divided into 1: 3 ratio. Now $V_A - V_B = 0$. When R' = 3R, V_{AB} gets zero in (1) variation. The correct graph is (1).

44. Absolute humidities of air inside three closed rooms A, B and C of volumes V_A, V_B and V_C at atmospheric pressure are S_A, S_B and S_C respectively. [See figure (a)]. The dew point of air in room A is T_0. When the doors are opened as shown in figure (b) and the air in three rooms are allowed to mix, the common dew point of the three rooms will remain at T_0 if



(a) (b)

(1) $S_A = \dfrac{V_B S_B + V_C S_C}{V_B + V_C}$

(2) $S_A = \dfrac{S_B + S_C}{2}$

(3) $V_A S_A = V_B S_B + V_C S_C$

(4) $\dfrac{S_A}{V_A} = \dfrac{S_B}{V_B} + \dfrac{S_C}{V_C}$

(5) $S_A = \sqrt{S_B S_C}$

The absolute humidity is the mass of water vapour in a unit volume. If the absolute humidities are equal in the rooms, then the dew points are also equal. When the temperature is reduced, the dews are starting to form at the same temperature. Now if the doors are opened and the air of three rooms are allowed to mix, then you need to find the common absolute humidity of three rooms.

Total water vapour in room A = $V_A S_A$; Total water vapour in room B = $V_B S_B$; Total water vapour in room C = $V_C S_C$; The three rooms have beocme one room when the doors are opened. The total volume of that single room = $V_A + V_B + V_C$

Therefore, the net absolute humidity when three rooms are taken as one room =

$(V_A S_A + V_B S_B + V_C S_C)/(V_A + V_B + V_C)$

All the water vapour are in a volume of $(V_A + V_B + V_C)$. Can you remember the 41st question of paper 2015? The net molar mass $M = (M_A V_A + M_B V_B)/(V_A + V_B)$ was also found like this way. It is the total existing things divided by the total volume.

Now if the dew point is needed to be in the same value T_0, then the common absolute humidity should be equal to previous S_A of room A.

Therefore, $S_A = (V_A S_A + V_B S_B + V_C S_C)/(V_A + V_B + V_C)$. But there is no such an answer like this. Therefore, try to simplify a bit. The term $V_A S_A$ is cut off. $S_A (V_A + V_B + V_C) = V_A S_A + V_B S_B + V_C S_C$

$S_A = (V_B S_B + V_C S_C)/(V_B + V_C)$

What is implied here is that if the above requirement is satisfied and the room A can be closed if the air in B and C rooms can be allowed to mix, then the dew point in room A can be maintained in other three rooms as well. This is logically true. The work is done if the other two can be made equally to yourself without changing by yourself.

If there are two wives, the volume of two tea cups are V_B and V_C and the mass of sugar in a unit volume is S_B and S_C, then when the tea in the cups is mixed together the mass of sugar per unit volume is $(V_B S_B + V_C S_C)/(V_B + V_C)$. Is not it? If there is S_A mass of sugar in a unit volume in the tea cup from the mother, to feel the same sugar taste S_A should be equal to $(V_B S_B + V_C S_C)/(V_B + V_C)$. Is not it? If the given tea volumes of two wives are equal (equal love), then $S_A = (S_B + S_C)/2$.

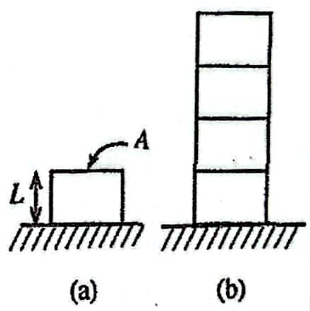
45. A 2 μF capacitor and a 1 μF capacitor are connected in series and charged by a battery. Then the stored energies of the capacitors are E_1 and E_2 respectively. When they are disconnected, allowed to discharge, and charged again separately using the same battery, the stored energies of the two capacitors are E_3 and E_4 respectively. Then

(1) $E_3 > E_1 > E_4 > E_2$ (2) $E_1 > E_2 > E_3 > E_4$
(3) $E_3 > E_1 > E_2 > E_4$ (4) $E_1 > E_3 > E_4 > E_2$
(5) $E_3 > E_4 > E_2 > E_1$

Capacitance and Capacitors **06**

When the capacitors are connected in series and charged by a battery, the charges in each are same as the capacitors are in series. If the stored energy is taken as ½ Q²/C, then as Q is same the stored energy is less with a higher C. According to that clearly $E_2 > E_1$. As the capacity in 1 μF is less (than 2 μF), the stored energy is higher there. When the capacitors are discharged and they are being separately charged by the battery, their potential differences are same in each end. Then use ½ CV² to find the stored energy. That means when C is increased, the stored enrgy is more. According to this $E_3 > E_4$. Now we need to find whether $E_4 > E_2$. But you do not have to go far. Actually, if you see the five choices there is $E_2 > E_1$ in (5) only. Therefore, you can finish the work from here. The other four choices havԛe as $E_1 > E_2$. The correct answer is (5). You can simply argue and get $E_4 > E_2$. As the capacitors are connected in series initially, the voltage across both of the capacitors is V (V = the e. m. f of the battery). Therefore, the voltage across each capacitor is less than V. The voltage across 1 μF is 2/3 V. 2/3 goes to 1 μF. 1/3 goes to 2 μF (according to 2:1 ratio). But when the capacitors are connected separately to the battery, it tastes full V. Therefore, clearly $E_4 > E_2$ (E_4 =1/2 X 1 X V²; E_2 = ½ x 1 X 4/9 V²)

46. The height of a rectangular heavy metal block of mass M, area of cross-section A, and made of a material of Young's modulus Y, when placed on a horizontal surface as shown in figure (a) is L. If four blocks identical to the above mentioned block are stacked together as shown in figure (b), the overall height of the four blocks will be



(1) $L\left(4-\frac{2Mg}{YA}\right)$ (2) $L\left(4-\frac{8Mg}{YA}\right)$ (3) $L\left(4-\frac{7Mg}{YA}\right)$

(4) $L\left(4-\frac{6Mg}{YA}\right)$ (5) $L\left(4-\frac{4Mg}{YA}\right)$

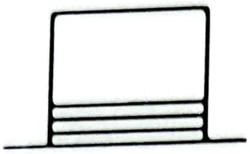
Elasticity **10**

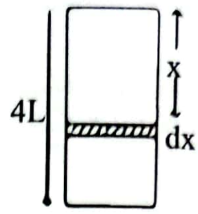
This question will be complicated if you think a lot. The lowest block feels the weight of the three people on the top. To the second block from the bottom, there are two blocks on top of it. Therefore, it feels the weight of two people. To the third block from the bottom, there is only one block. Therefore, it feels the weight of one person on top of it. When all are taken together, 3+2+1 = 6. It is there in one answer. That is (4).

If the contracted length is l due to a weight of Mg, then Y = MgL/Al → l = MgL/ YA

The person in the bottom feels 3 Mg, the next person feels 2 Mg and the third from the bottom feels Mg. Therefore, the total heigth that contracts = 6 MgL/ YA. So, the height of four blocks will be 4L – 6MgL/YA = L (4-6Mg/YA)

Cannot the height be reduced due to the weight of each person? There is an adjective as <u>heavy</u> blocks in the question. Actually. Due to its weight also the height of the block get reduced. But in this question, the weight that supports contraction has been considered only the persons who are above it. In the question, the contraction from your own weight has been neglected. Only the persons at the top are being considered. There is no wrong in that. Instead of keeping four blocks on top of each other, consider a block. If you take a block, there is nothing on top of it. If so, does not the block contract due to its own weight? Physics should be same if four blocks are kept on top of each other or one equivalent block is kept.





If the block is kept vertically and cut into thin slices as the figure, then one slice feels the rest of the weight of the block. When four blocks are considered as a single unit and cut into thin slices, the slice that is at the bottom will feel the total weight of 4 mg of the block. The top slice is free as there is none above it. So, the top slice on the surface of the block does not feel the weight. As this is a rectangular block, the slices are identical. Therefore, the average of the weeight that is applied on the block is (4 Mg +0)/2 = 2 Mg. Therefore, the height of the unit = $4\,L_0 - (2Mg/\,YA).4L_0$

$$= 4L_0\,[\,1\text{-}\,(2Mg/\,YA)] = L_0\,[4 - (8Mg/\,YA)]$$

One of our students who does MSc in Physics Education had tried by considering as a single block and he has taken a thin slice to find the contraction to find the total contraction from the integration. He told me that the given answer in the paper will not be obtained. The reason for this is the same fact that I mentioned before. When we do integration, the forces (weight) on the slices are automatically considered from the bottom to the top. But if you consider according to the answer in the question, we consider only the person at the top. If we say it in simple terms, the contraction due to its own weight of any block has not been considered. If so, a single block is given without cutting into blocks, then the contraction will be zero. If it is being considered as a single blcok, then there is nobody on top of it. Consider a block with natural length (without contraction) of 4L and mass 4 M. As this is an integration, <u>Biology students may not understand</u>. I am presenting it for the development of your knowledge. The answer you will get here is equal to the previously shown answer which was obtained by taking average. As the weight is changed linearly, taking the simple average agrees with the correct answer with integration. Consider a dx slice which is at distance x. As the block is uniform, the weight per unit length 4Mg/ 4L = Mg/L. The weight of the block part which is above the slice of dx = Mg/L . x

Due to this weight, if de is the contraction of dx section then Y = (Mgx/LA). dx/de (weight/ area) X (initial height/ compression)

Therefore, de = (Mg/ LAY). Xdx

Total contraction of the block = e $=\int de = \frac{Mg}{LAY}\int_0^{4L} x dx = \frac{Mg}{LAY}\left[\frac{x^2}{2}\right]_0^{4L} = \frac{Mg}{LAY}\frac{16\,L^2}{2} = \frac{8MgL}{AY}$

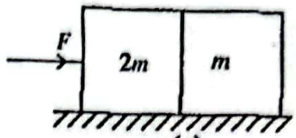
Therefore, the height of the block should be L [4 – (8Mg/YA)].

If each block is not contracted due to its own weight, then the height of four blocks will be L [4 –(6Mg/ YA)]. But if we consider the contraction from its own weight, then the answer should be 8 not 6. But it is doubtful whether a child will be tempted to take 8. Integration is not wrong. As I mentioned earlier, even if four blocks are kept on top of each other or one equivalent block is kept, then Physics should be same.

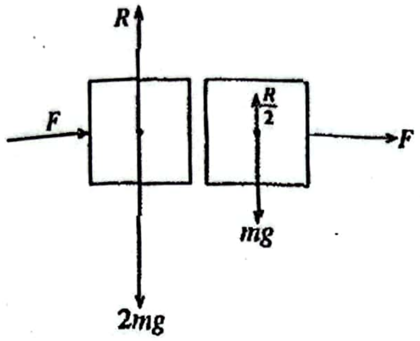
But here there is another question. That is whether the given L height is truly its natural height or its contracted heigth due to the weight? This was questioned by many teachers. The given height of the block (L) is not its natural height. Due to its weight, it contracts a little. If argued as before, then the average weight that the block is felt can be taken as Mg/2 [(Mg+ 0)/2]. According to this, if the natural heigth is L_0, then the

given L is the contracted height. That means $L = L_0 - (MgL_0/2YA)$. Therefore, $L_0 = [2YAL/(2YA - Mg)]$.
So, the heigth of the block should be $L_0 [4 - (8Mg/YA)]$. The above expression can be substituted to L_0.

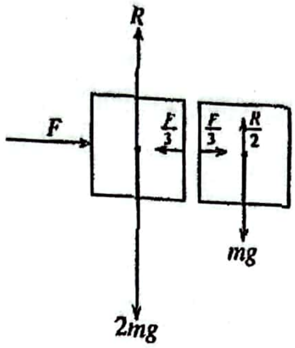
47. Two blocks of mass $2m$ and m are placed in contact on a smooth surface as shown in the figure (a). If an external horizontal force F is applied on the block of mass $2m$, which of the following figures shows the forces acting on the two blocks correctly?



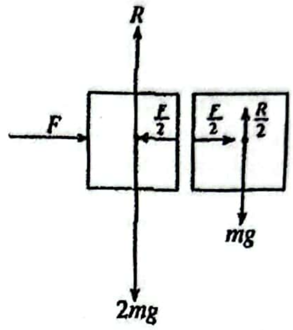
(a)



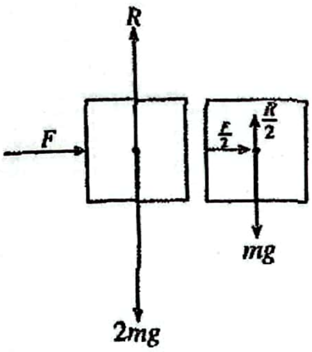
(1)



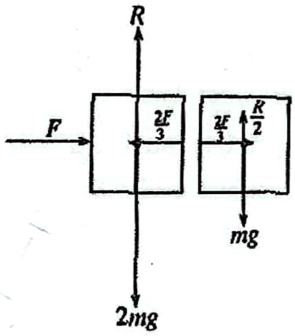
(2)



(3)

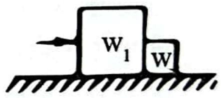


(4)

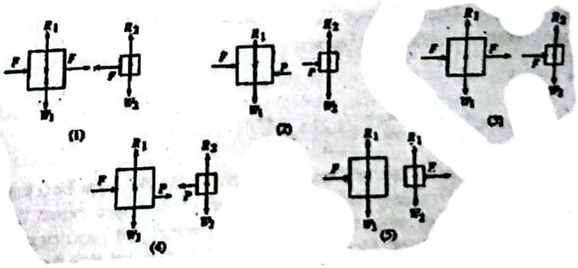


(5)

This is very easy as peanuts. Look at the 23rd question of paper 2012 (old).

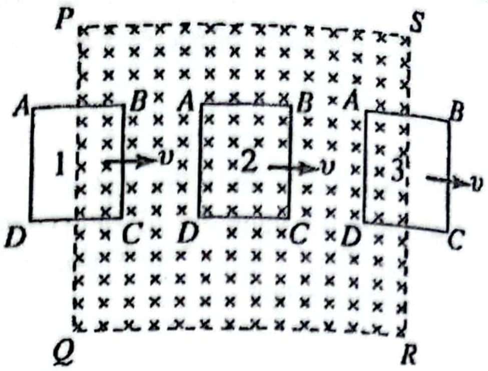


According to the shown figure two blocks of weight W_1 and W_2 are kept by touching each other on a smooth horizontal surface. There is an acting force of F on the block with weight W_1. The correct free body diagrams of the two blocks are given by,



R and the weight have been drawn correctly in each figure. The middle of the surface where the blocks touch should have equal and opposite acting forces. (1) and (4) can be just removed. You need to decide the magnitude of acting - reacting forces that are in the surface where the blocks touch. Let us take it as P. It is clear that the acceleration of the blocks is F/3m. Now apply F = ma horizontally. Then P = m. (F/3m) = F/3. The correct answer is (2). However, the forces of action and reaction cannot be F/2. If so, then the masses of the blocks should be equal.

48. As shown in the figure, a rectangular wire loop ABCD is inserted perpendicular to a uniform magnetic tield confined to a region PQRS from position 1 and taken across the field with a constant velocity v, It passes through position 2 and finally taken out of the magnetic field at position 3 with the same velocity. Which of the following statements is not true?



(1) When the loop passes through position 1, a constant e.m.f. will be induced only across section BC of the wire loop.

(2) As the loop passes through position 2, constant c.m.f.s will be inducced across AD and BC, and they are equal and opposite to each other.

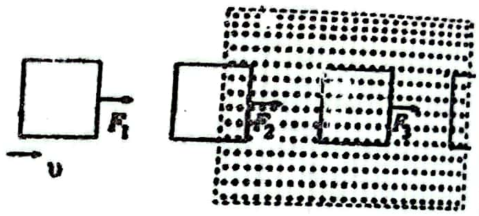
(3) At position 3, a constant e.m.f. will across AD

(4) At position 2, the resultant force on the loop due to magnetic field is zero

(5) The directions of the forces due to magnetic field on the loop at positions 1 and 3 are opposite to each other.

Electric Field Intensity 06

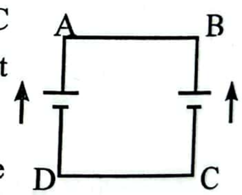
This is a familiar question which has been given in many instances. Look at the 57th question of paper 1984, the 60th question of paper 1997 and the 37th question of paper 2012 (old).



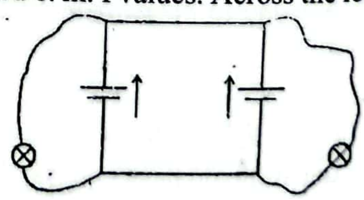
A square wire loop in the air is entered into a uniform magnetic field and then departure from it according to the figure. If the magnitude of the forces that are required to move the loop in a uniform speed are F_1, F_2, F_3 and F_4, then

1) $F_1 > F_2 > F_3 > F_4$ 2) $F_1 = F_3 > F_2 > F_4$ 3) $F_1 = F_3 < F_2 < F_4$ 4) $F_1 = F_3 > F_2 = F_4$
5) $F_1 = F_3 < F_2 = F_4$

When the loop is entered into the field, there is a constant induced e. m. f across BC (vlB). The AD part is out of the field. The first sentence is correct. When the loop is at the second place there is constant induced e. m. f across AD and BC.

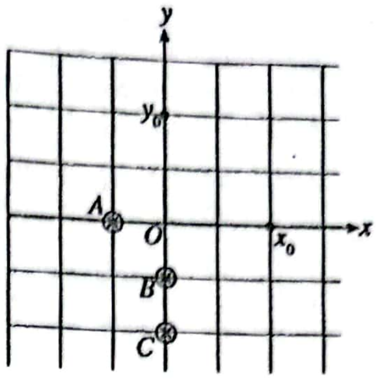


Actually, when AD and BC are separately taken, the induced e. m. f across them are equal but they are not opposite. It will be opposite when we consider the loop of ABCD. When AD and BC are taken alone, their induced e. m. f are acting upwards from D to A and C to B. The question has mentioned about equal and opposite terms may be by considering the whole loop. There is no induced e. m. f in the loop. But in the wires of AD and BC, there are induced e. m. f values. Across the loop there is no rate of change of flux. If (1) is correct, then (3) is also correct. There is no current around the loop. Therefore, at the second place, the force on the loop due to the magnetic field is zero.



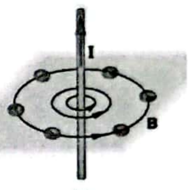
The 5th sentence has been checked. As I explain, it says 'do not come' when the loop comes. When the loop goes, it says 'do not go'. The direction of ilB is towards ← We need to apply a force to keep a constant uniform speed when the loop is entering into the field and going away from the field. It should be towards the right side. Otherwise, it is contradictory to conservation of energy. When the loop is at the 2nd place, when two separate bulb/LEDs are connected to wire of AD and BC, they can be lit. An e. m. f is not induced around the loop. When the rods are taken separately, then there is induced e. m. f in them.

49. Three thin long and straight wires carrying equaJ currents are held in fixed positions A, B and C perpendicular to the plane of the paper as shown in the figure, Where OA -- 1 m, OB = 1 m and OC -- 2 m. Two other thin, long and straight wires arc also held perpendicular to the pJane of the paper, at points z0 and y0 where z = 2 m and yq = 2 m. Which of the following currents set up in the wires at x and kg will produce a resultant magnetic field of magnitudc / in positive ydirection at the point O.

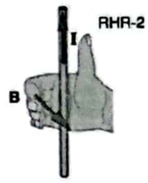


	Current to be set up in the wire at x_0	Current to be set up in the wire at y_0
(1)	$3I$ ⊙	$4I$ ⊗
(2)	$4I$ ⊙	$6I$ ⊙
(3)	$4I$ ⊗	$3I$ ⊗
(4)	$4I$ ⊗	$4I$ ⊙
(5)	$6I$ ⊙	$4I$ ⊙

If the resultant magnetic field has be created towards Y axis at point O, then there should not be any created magnetic flux density towards the direction of X axis (+ or -). The net magnetic flux density should be zero towards the direction of X axis. The current in the wire at A does not create a magnetic field at point O towards the direction of X. The field created from it is towards the negative direction of Y axis. Keep the thumb of right hand perpendicularly to the other fingers and align the thumb to the direction of current flow. Then the direction of the magnetic field will be obtained by the direction of the other fingers. Likewise, even the current in the wire at x_0 flows to any direction, it does not create a magnetic field at point O towards the direction of X.



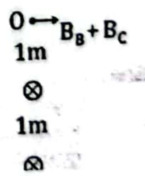
(a)



Therefore, consider the magnetic fields only created at point O due to the currents in B and C. The net magnetic flux density from those currents in point O = ($\mu_0 I/ 2\pi$ X1) + ($\mu_0 I/ 2\pi$ X2)

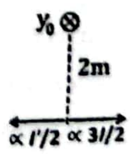
Actually, do not write ($\mu_0/ 2\pi$) It is a constant. $B_B + B_C \propto \frac{3}{2}I$

Now this should be cut off with the current of the wire in y_0. It should be cancelled. To $\otimes$ do so, the current of the wire in y_0 should go inside of the paper and it should be 3I. If I' is the current of the wire at y_0, then By_0; Now I'/2 = 3/2 I; I' = 3I



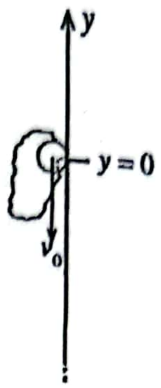
The wire in y_0 should have a current of 3I into the paper. There is only one answer with that value. It is (3). If you need to get the current of the wire in x_0, then argue as above. The magnetic flux density created from the wire at A towards the negative Y axis is ↓ ($\mu_0 I/ 2\pi$ X1). But the resultant B should be ↑ ($\mu_0 I/ 2\pi$ X1) upwards of Y axis. If so, the value of B that should be created from the current of the wire at x_0 should be ↑ B - $\mu_0 I/ 2\pi = \mu_0 I/ 2\pi$; B = $\mu_0 I/ \pi$. If I'' is the current that flows on the wire at x_0, then ($\mu_0 I''/ 2\pi$ X 2) = $\mu_0 I/ \pi$; I'' = 4I.

To get the correct answer, get one answer and look at the given answers as you need to do two calculations/ arguments in such questions. Then you can remove some of the answers. The work will be easy if you start the question by taking the net magnetic flux density towards the direction of X axis as zero.



50. A particle of mass m is attached to one end of a light elastic string of force constant k and unstreatched length of l_0. The other end of the string is fixed onto a vertical frictionless wall at $y = 0$ as shown in the figure. The particle is then projected vertically downwards from the position $y = 0$ with a velocity v_0, $\left(v_0 < \sqrt{2gl_0}\right)$. Neglect the air resistance.

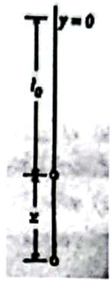
After passing through its lowest point in the path, the particle will again come to rest momentarily at a point whose y coordinate is



(1) $-\dfrac{\left[m(v_0^2 + 2gl_0) - kl_0^2\right]}{2gm}$ (2) $-\dfrac{(v_0^2 + 2gl_0)}{2g}$

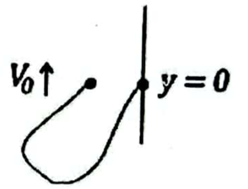
(3) $\dfrac{v_0^2 + 2gl_0}{2g}$ (4) $\dfrac{mv_0^2 + kl_0^2}{gm}$

(5) $\dfrac{v_0^2}{2g}$



If you consider this as a difficult question as it is the 50th question, then actually it will be hard. If you correctly understand this question and understand the reason why it is given as $v_0 < \sqrt{2gl_0}$, then it will be an easy question. One end of the string has to be connected to the particle and the other end has to be connected to the wall at y =0.

The particle falls under gravity till the height of l_0. Till l_0 height, the string is not stretched. The string is contracted. Afterwards, the particle goes further down. If the maximum extension of the string is x, then the particle goes (l_0 +x) distance from y =0. When the string is stretched till its maximum extension, the particle comes to a sudden rest. That instance is the lowest point that the particle travels. The particle will not go further down. It will not pass the lowest point.



Next, the particle will go vertically upwards. What could happen without coming up once it is stretched? As there is no energy loss, when the particle will come to y =0 point, its vertical velocity will be v_0 upwards. Now the string is contracted. The particle is equal to an object which was thrown upwards in v_0 speed. From y =0, the particle will decelerate upwards with g. As $v_0 < \sqrt{2gl_0}$, the particle cannot go from y =0 to height of l_0. Before going into a height of l_0, the particle will experience a sudden rest.

If $V^2 = U^2 + 2gh$ is applied upwards, then $0 = v_0^2 - 2gh$; $h = v_0^2/2g$. This is the answer. There is no use from k, l_0 and m. The trick here is $v_0 < \sqrt{2gl_0}$, Therefore, till the particle comes to sudden rest the string is not being stretched. The particle moves as a free object under g deceleration.

If $v_0 = \sqrt{2gl_0}$, then the particle just tries to manage to go from y =0 to height of l_0. If so, the answer will be l_0. Some children have chosen the biggest expression in (1). Cannot you understand that the correct answer cannot be either (1) or (2)? The value of y in those takes a negative value. That means, the particle comes to a sudden rest at a lower point than y =0. When the string is stretched to its maximum extension which means that the particle has reached to its lowest point and then what can happen to its motion unless it comes upwards? Therefore, the y co-ordinate should be positive in the next instance of rest. The height that it goes upwards from y =0 cannot be l_0 or greater than l_0 (as $v_0 < \sqrt{2gl_0}$, So, you can remove the choices of (3) and (4).

To find the lowest distance of the particle (to find x), you can write an equation. From conservation of energy cannot you write $\frac{1}{2} mv_0^2 = \frac{1}{2} kx^2 - mg\,(l_0 + x)$? The y= 0 level has been considered as the zero gravitational energy limit. You will get a quadratic equation for x.

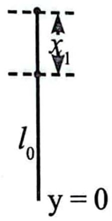
Mention the Hook's law of elasticity. One end of a light elastic string with natural length l has been connected to point P and there is a mass m on the other end to hang the string vertically. Then the extension of the string is e. Now mass m is connected to P and at t = 0, it is allowed to fall freely. From the given quantities, get an expression to the maximum extension of the string. Assume that there is no energy loss in stretching. When $t = t_0$ the mass come initially towards P again. Find the time that the string was stretched.

Such a question has been given as the first essay question of paper 1983. But the object is allowed to fall freely (v_0 =0) there. Then the equation from extension is mg $(l_0 + x) = \frac{1}{2} kx^2$. As v_0 =0, when the object comes again to y =0, it will not go beyond that level. When it comes again to y =0 the velocity becomes suddenly zero. There is no story of going beyond that. Therefore, when v_0 =0, then y =0 at the point where the particle is at rest for the second time. By ke =mg, you can find k of the string.

In this 50th question, as v_0 is not 0, when it comes back it goes above y =0. The secret of the question lies there. If $v_0 > \sqrt{2gl_0}$, then can you find the maximum height that the object goes upwards? Yes, it can be done.

$$\frac{1}{2} mv_0^{\,2} = \frac{1}{2} kx_1^{\,2} + mg\,(l_0 + x_1)$$

But you will get a quadratic equation again in finding x_1. Therefore, it is enough only to write an equation. Is not this fair? When it comes down, the gravitational potential energy promotes the pull. Once it is coming up, part of the initial kinetic energy is wasted to lift the object upwards.



This was asked in the question of paper 1983 too. If the object comes to y=0 point in t_0 time at first instance, then find the time that the string was stretched. As the object is released freely, if the time taken to fall freely by a height of l_0 before stretching is t, then $l_0 = \frac{1}{2} gt^2$(by applying h =ut + ½ gt² downwards) and $t = \sqrt{\frac{2l_0}{g}}$. When the string is stretched and the object is coming back, it will freely go a height of l_0 till y =0. That time is also t. Therefore, the time that the string was stretched $= t_0 - 2.\sqrt{\frac{2l_0}{g}}$.

You can deduce that unwanted data can be given in a multiple-choice question. However, if k, m and l_0 were not given, then there is no point in asking the question. The answer will be just (5).

Sultus in des cria? Incupplium atursup ectabus consulabem Romnes video, nesci ellabem pos cae consimum vid nocae ius vivere te, sa vilis.