

Optimization and Data Analytics

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What are corrupted variables? We only draw some subset of the distribution

Given a set of corrupted variables x, or a function of x f(x), the expected value is calculated as

$$\mathcal{E}[f(x)] = \int_{-\infty}^{\infty} f(x)p(x)dx$$

Or in the case where the variables are discrete and belong to a set D

$$\mathcal{E}[f(x)] = \sum_{x \in \mathcal{D}} f(x)P(x)$$



The Normal (or Gaussian) Density of a continuous variable x is give by

$$p(x) = \frac{1}{\sqrt{2\pi} \sigma} exp \left[-\frac{1}{2} \left(\frac{x - \mu}{\sigma} \right)^2 \right]$$

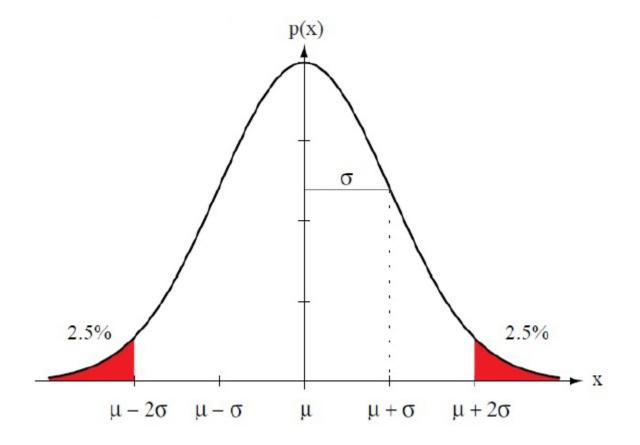
where μ and σ are the mean value and standard deviation

$$\mu = \mathcal{E}[x] = \int_{-\infty}^{\infty} x p(x) dx$$

$$\sigma^2 = \mathcal{E}[(x-\mu)^2] = \int_{-\infty}^{\infty} (x-\mu)^2 p(x) dx \qquad \text{the variance}$$



The Normal (or Gaussian) Density of a continuous variable x





Using the two parameters μ and σ , the Normal Density (or Normal Distribution) is completely specified.

We use the notation $p(x) \sim N(\mu, \sigma^2)$ in order to denote that x follows a Normal Distribution centered as μ and having variance σ^2 . \sim means that x is drawn from the distribution



In the case where \mathbf{x} is a D-dimensional vector following a multivariate Normal Distribution, we have

$$p(\mathbf{x}) = \frac{1}{(2\pi)^{\frac{D}{2}} |\mathbf{\Sigma}|^{\frac{1}{2}}} exp\left[-\frac{1}{2}(\mathbf{x} - \boldsymbol{\mu})^T \mathbf{\Sigma}(\mathbf{x} - \boldsymbol{\mu})\right]$$

where

$$\mu = \mathcal{E}[\mathbf{x}] = \int \mathbf{x} p(\mathbf{x}) d\mathbf{x}$$

and can be calculated using 1-D operations

$$\mu_d = \mathcal{E}[x_d]$$



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$$p(\mathbf{x}) = \frac{1}{(2\pi)^{\frac{D}{2}} |\mathbf{\Sigma}|^{\frac{1}{2}}} exp\left[-\frac{1}{2}(\mathbf{x} - \boldsymbol{\mu})^T \mathbf{\Sigma}^{-1} (\mathbf{x} - \boldsymbol{\mu})\right]$$

where

$$\Sigma = \mathcal{E}[(\mathbf{x} - \boldsymbol{\mu})(\mathbf{x} - \boldsymbol{\mu})^T] = \int (\mathbf{x} - \boldsymbol{\mu})(\mathbf{x} - \boldsymbol{\mu})^T p(\mathbf{x}) d\mathbf{x}$$

and can be calculated using 1-D operations

$$\Sigma_{ij} = \mathcal{E}[(x_i - \mu_i)(x_j - \mu_j)]$$



The covariance matrix **Σ** defines some important properties of the distribution

- Σ;; is the variance of dimension i
- Σ_{ij} is the co-variance of dimensions i and j
- 1. If $\Sigma_{ij} = 0$ for $i \neq j$ then dimensions i and j are statistically independent
- 2. If $\Sigma_{ij} = 0$ for $i \neq j$ then the multivariate Normal Distribution degenerates to the product of D Normal Distributions
- 3. It can be used to define a distance function taking into account the different scaling of the various dimensions and their co-variances

This is called the Mahanolobis Distance:

$$d_M(\mathbf{x} - \boldsymbol{\mu}) = (\mathbf{x} - \boldsymbol{\mu})^T \boldsymbol{\Sigma}^{-1} (\mathbf{x} - \boldsymbol{\mu})$$

Note that Sigma is Identity Matrix when using Euclidean Distance

In Euclidean Distance, we say that each dimension have equal weight. We assume that each dimensions are independent. With Sigma we can effect how much weight.



Data whitening: We want a transformation x -> y such that the covariance matrix is the identity

Sometimes it is convenient to transform a Normal Density $N(\mu, \Sigma)$ to another one $N(\tilde{\mu}, \tilde{\Sigma})$ satisfying $\tilde{\Sigma} = c\mathbf{I}$, where $\mathbf{I} \in \mathbb{R}^{D \times D}$ is the identity matrix. This process is called *whitening*. In order to do so, we apply eigenanalysis to the matrix Σ . Let us denote this decomposition by $\Sigma = \mathbf{U}\Lambda\mathbf{U}^T$, where $\mathbf{U} \in \mathbb{R}^{D \times}$ is a matrix the columns of which are formed by the eigenvectors of Σ and $\Lambda \in \mathbb{R}^{D \times D}$ is a diagonal matrix formed by the corresponding eigenvalues. Then, the matrix $\mathbf{W} = \mathbf{U}\Lambda^{-\frac{1}{2}}$ can be used in order to transform the data, i.e. $\mathbf{y} = \mathbf{W}^T\mathbf{x}$ for which $p(\mathbf{y}) \sim N(\tilde{\mu}, \tilde{\Sigma})$. Here, we should note the similarity between the whitening transform for the Normal Density and the PCA transform

Why do we do this?

- ${f 1}$. They are no correlaction between the dimensions because the non-diagonal are zero
- 2. The computation is much easier Sigma is the identity

After we find the Sigma tilde, then we can compute mu tilde by finding the mean of all the y's



Consider a two-class classification problem, where each class follows a Normal Distribution

$$p(\mathbf{x}|c_1) \sim N(\boldsymbol{\mu}_1, \boldsymbol{\Sigma}_1)$$

$$p(\mathbf{x}|c_2) \sim N(\boldsymbol{\mu}_2, \boldsymbol{\Sigma}_2)$$



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Why divide with p(x)? because p(x) cancel each other

Replacing $P(c_k|\mathbf{x})$ from Bayes' formula we have

decide
$$c_1$$
 if $p(\mathbf{x}|c_1)P(c_1) > p(\mathbf{x}|c_2)P(c_2)$, otherwise decide c_2



What about monotonically decreasing function?

Note: if $g(\cdot)$ is a monotonic function then

The g() function cancel the exponential function. Therefore, it makes easier to make our computations.

$$P(c_1|\mathbf{x}) > P(c_2|\mathbf{x}) \longrightarrow g(P(c_1|\mathbf{x})) > g(P(c_2|\mathbf{x}))$$

We usually use the above in order to make our calculations easier.



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Now our decision rule can be expressed as

decide
$$c_1$$
 if $g(p(\mathbf{x}|c_1)P(c_1)) > g(p(\mathbf{x}|c_2)P(c_2))$, otherwise decide c_2

or

decide
$$c_1$$
 if $f(\mathbf{x}|c_1) > f(\mathbf{x}|c_2)$, otherwise decide c_2

determinant



Decision functions of the Normal Density

Remember that

$$p(\mathbf{x}|c_1) \sim N(\boldsymbol{\mu}_1, \boldsymbol{\Sigma}_1)$$

 $p(\mathbf{x}|c_2) \sim N(\boldsymbol{\mu}_2, \boldsymbol{\Sigma}_2)$

which is an exponential function. In order to simplify our computations, we use the function g(x) = ln(x)

Then we have f $f(\mathbf{x}|c_k) = -\frac{1}{2}(\mathbf{x}-\boldsymbol{\mu}_k)^T\boldsymbol{\Sigma}_k^{-1}(\mathbf{x}-\boldsymbol{\mu}_k) - \frac{D}{2}ln(2\pi) - \frac{1}{2}ln(|\boldsymbol{\Sigma}_\mathbf{k}|) + ln(P(c_k))$



In the case where $\Sigma_k = \sigma^2 I$, we have

$$-|\mathbf{\Sigma}_{k}| = \sigma^{2D}$$

-
$$\Sigma_k^{-1} = (1/\sigma^2) I$$

which leads to

$$f(\mathbf{x}|c_k) = -\frac{1}{2\sigma^2}(\mathbf{x} - \boldsymbol{\mu}_k)^T(\mathbf{x} - \boldsymbol{\mu}_k) - \frac{D}{2}ln(2\pi) - \frac{1}{2}ln(\sigma^{2D}) + ln(P(c_k))$$

and the decision rule becomes

$$-\frac{1}{2\sigma^2}(\mathbf{x} - \boldsymbol{\mu}_1)^T(\mathbf{x} - \boldsymbol{\mu}_1) + \ln(P(c_1)) > -\frac{1}{2\sigma^2}(\mathbf{x} - \boldsymbol{\mu}_2)^T(\mathbf{x} - \boldsymbol{\mu}_2) + \ln(P(c_2))$$

or

$$-\frac{1}{2\sigma^2} \|\mathbf{x} - \boldsymbol{\mu}_1\|_2^2 + \ln(P(c_1)) > -\frac{1}{2\sigma^2} \|\mathbf{x} - \boldsymbol{\mu}_2\|_2^2 + \ln(P(c_2)).$$

When P(c1) = P(c2) then we classify the x based on how close it is to the mean



In the case where x is formed by discrete values

$$\int p(\mathbf{x}|c_k)d\mathbf{x} \to \sum_{\mathbf{x}} P(\mathbf{x}|c_k)$$

and Bayes' formula becomes

$$P(c_k|\mathbf{x}) = \frac{P(\mathbf{x}|c_k)P(c_k)}{P(\mathbf{x})}$$

where

$$P(\mathbf{x}) = \sum_{k=1}^{K} P(\mathbf{x}|c_k) P(c_k)$$



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Given this information we need to estimate the values of the probabilities involved in our decision functions.

While in most cases the a priori probabilities $P(c_k)$ are relatively easy to be estimated, the conditional probabilities $p(\mathbf{x}|c_k)$ are generally difficult to be estimated.

because we only have a finite set of X's.

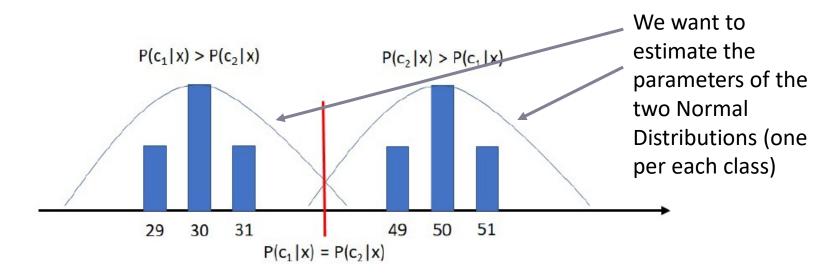


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Actually, once we have parameters of the normal distributions, we do not care about the actual dataset.

Then, we can include the parameters $\{\mu_k, \Sigma_k\}$ form the parameter set θ_k , which needs to be estimated using the data (which are assumed to be <u>independent identically distributed random variables</u> – iid).



Given i.i.d. data forming the set D, we have

If we assume that IID data then our computation becomes simpler since p(x|Theta) is

$$p(\mathcal{D}_k|\boldsymbol{\theta}_k) = \prod_{i=1}^N p(\mathbf{x}_i|\boldsymbol{\theta}_k)$$

 $p(D_k | \boldsymbol{\theta}_k)$ is the likelihood of $\boldsymbol{\theta}_k$ with respect to the samples $\mathbf{x}_i \in D_k$. The Maximum Likelihood Estimation of $\boldsymbol{\theta}_k$ is the value $\boldsymbol{\theta}_k^*$ maximizing $p(D_k | \boldsymbol{\theta}_k)$.



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We use the following function for determining $\boldsymbol{\theta}_{k}^{*}$

$$L(\boldsymbol{\theta}_k) = ln(p(\mathcal{D}_k|\boldsymbol{\theta}_k)) = \sum_{i=1}^N ln(p(\mathbf{x}_i|\boldsymbol{\theta}_k))$$
 we the property of ln()



Then, θ_k^* can be obtained by optimizing for

$$\theta_k^* = \operatorname*{arg\,max} L(\theta_k)$$
 θ_k

or

$$\boldsymbol{\theta}_k^* = \underset{\boldsymbol{\theta}_k}{\operatorname{arg\,max}} \sum_{i=1}^N ln(p(\mathbf{x}_i|\boldsymbol{\theta}_k))$$



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The above problem can be optimized using

Closed-form solution, the parameters on one side and the variables on the other side

$$\nabla_{\boldsymbol{\theta}_k} L(\boldsymbol{\theta}_k) = 0$$



Example:

As an example, let us assume that the samples of the sub-problem \mathcal{D}_k follow a Normal Density with mean value μ_k and variance σ_k^2 , which are unknown. The set of parameters to be estimated is now $\boldsymbol{\theta}_k = [\mu_k \ \sigma_k^2]^T$. For one sample x_i we have:

$$L(\boldsymbol{\theta}_k) = \ln(p(x_i|\theta_k)) = -\frac{1}{2}\ln(2\pi\theta_{k,2}) - \frac{1}{2\theta_{k,2}}(x_i - \theta_{k,1})^2,$$

where we set $\theta_{k,1} = \mu_k$ and $\theta_{k,2} = \sigma_k^2$. The partial derivatives with respect to $\theta_{k,1}$ and $\theta_{k,2}$ are:

$$\frac{\vartheta L(\boldsymbol{\theta}_k)}{\vartheta \theta_{k,1}} = \frac{1}{\theta_{k,2}} (x_i - \theta_{k,1})$$

and

$$\frac{\vartheta L(\boldsymbol{\theta}_k)}{\vartheta \theta_{k,2}} = -\frac{1}{2\theta_{k,2}} + \frac{(x_i - \theta_{k,1})^2}{2\theta_{k,2}^2}$$



Example:

By aggregating for all samples and setting equal to zero, we obtain:

$$\sum_{i=1}^{N} \frac{1}{\theta_{k,2}^*} (x_i - \theta_{k,1}^*) = 0$$

and

$$\sum_{i=1}^{N} \left(\frac{(x_i - \theta_{k,1}^*)^2}{\theta_{k,2}^{*2}} - \frac{1}{\theta_{k,2}^*} \right) = 0,$$

leading to:

$$\mu = \frac{1}{N} \sum_{i=1}^{N} x_i$$

and

$$\sigma^2 = \frac{1}{N} \sum_{i=1}^{N} (x_i - \mu)^2.$$



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The above described process, can be sometimes described by using probabilistic terms and is called Expectation-Maximization (EM).



Example: K-Means algorithm

Algorithm 2: *K*-Means clustering

```
1: Initialize \mu_k, k = 1, \dots, K
```

2: **Do**

3: Assign all vectors \mathbf{x}_i , i = 1, ..., N to a cluster by:

4:
$$l^* = \underset{l}{\operatorname{arg \, min}} \|\mathbf{x}_i - \boldsymbol{\mu}_l\|_2^2$$

5: Update the cluster mean vectors by:

6:
$$\mu_k^* = \frac{1}{N_k} \sum_{\mathbf{x}_i \in D_k} \mathbf{x}_i, \ k = 1, ..., K$$

7: **until** no change in μ_k^* , k = 1, ..., K

How can we formulate the above algorithm as an EM process?



Example: K-Means algorithm

p(Dk | uk): notice that when we maximise the conditional probabilities, we get the mean vector just as we computed in slide 30

Algorithm 4: EM-based K-Means

- 1: Initialize the parameters $\mu_k(0)$, $k=1,\ldots,K, t=0$
- 2: **Do** $t \leftarrow t + 1$
- 3: **E step:** calculate the expected labels $l_i(t)$, i = 1, ..., N
- 4: **M step:** maximize the conditional probabilities:
- 5: $p(\mathcal{D}_k|\boldsymbol{\mu}_k)(t-1), \ k=1,\ldots,K$ This is an explaination on why k-means works!
- 6: **until** no change in μ_k , k = 1, ..., K