

Python Machine Learning Equation Reference

Sebastian Raschka
`mail@sebastianraschka.com`

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Code Repository and Resources::
<https://github.com/rasbt/python-machine-learning-book>

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Chapter 1

Giving Computers the Ability to Learn from Data

- 1.1 Building intelligent machines to transform data into knowledge
- 1.2 The three different types of machine learning
- 1.3 Making predictions about the future with supervised learning
 - 1.3.1 Classification for predicting class labels
 - 1.3.2 Regression for predicting continuous outcomes
- 1.4 Solving interactive problems with reinforcement learning
- 1.5 Discovering hidden structures with unsupervised learning
 - 1.5.1 Finding subgroups with clustering
 - 1.5.2 Dimensionality reduction for data compression
- 1.6 An introduction to the basic terminology and notations

The Iris dataset, consisting of 150 samples and 4 features, can then be written as a 150×4 matrix $\mathbf{X} \in \mathbb{R}^{150 \times 4}$:

$$\begin{bmatrix} x_1^{(1)} & x_2^{(1)} & x_3^{(1)} & \dots & x_4^{(1)} \\ x_1^{(2)} & x_2^{(2)} & x_3^{(2)} & \dots & x_4^{(2)} \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ x_1^{(150)} & x_2^{(150)} & x_3^{(150)} & \dots & x_4^{(150)} \end{bmatrix}$$

For the rest of this book, unless noted otherwise, we will use the superscript (i) to refer to the i th training sample, and the subscript j to refer to the j th dimension of the training dataset.

We use lower-case, bold-face letters to refer to vectors ($\mathbf{x} \in \mathbb{R}^{n \times 1}$) and upper-case, bold-face letters to refer to matrices, respectively ($\mathbf{X} \in \mathbb{R}^{n \times m}$), where n refers to the number of rows, and m refers to the number of columns, respectively. To refer to single elements in a vector or matrix, we write the letters in italics $x^{(n)}$ or $x_m^{(n)}$, respectively. For example, x_1^{150} refers to the first dimension of the flower sample 150, the sepal length. Thus, each row in this feature matrix represents one flower instance and can be written as four-dimensional row vector $\mathbf{x}^{(i)} \in \mathbb{R}^{1 \times 4}$

$$\mathbf{x}^{(i)} = \begin{bmatrix} x_1^{(i)} & x_2^{(i)} & x_3^{(i)} & x_4^{(i)} \end{bmatrix}.$$

Each feature dimension is a 150-dimensional column vector $\mathbf{x}_j \in \mathbb{R}^{150 \times 1}$, for example

$$\mathbf{x}_j = \begin{bmatrix} x_j^{(1)} \\ x_j^{(2)} \\ \vdots \\ x_j^{(150)} \end{bmatrix}.$$

Similarly, we store the target variables (here: class labels) as a 150-dimensional column vector

$$\mathbf{y} = \begin{bmatrix} y^{(1)} \\ y^{(2)} \\ \vdots \\ y^{(150)} \end{bmatrix}, (y \in \{\text{Setosa, Versicolor, Virginica}\}).$$

1.7 A roadmap for building machine learning systems

1.7.1 Preprocessing – getting data into shape

1.7.2 Training and selecting a predictive model

1.7.3 Evaluating models and predicting unseen data instances

1.8 Using Python for machine learning

1.8.1 Installing Python packages

1.9 Summary

Chapter 2

Training Machine Learning Algorithms for Classification

2.1 Artificial neurons – a brief glimpse into the early history of machine learning

We can then define an activation function $\phi(z)$ that takes a linear combination of certain input values \mathbf{x} and a corresponding weight vector \mathbf{w} where z is the so-called net input ($z = w_1x_1 + \dots + w_mx_m$):

$$\mathbf{w} = \begin{bmatrix} w^{(1)} \\ w^{(2)} \\ \vdots \\ w^{(m)} \end{bmatrix}, \mathbf{x} = \begin{bmatrix} x^{(1)} \\ x^{(2)} \\ \vdots \\ x^{(m)} \end{bmatrix}.$$

Now, if the activation of a particular sample $x^{(i)}$, that is, the output of $\phi(z)$, is greater than a defined threshold θ , we predict class 1 and class -1, otherwise. In the perceptron algorithm, the activation function $\phi(\cdot)$ is a simple *unit step function*, which is sometimes also called the *Heaviside step function*:

$$\phi(z) = \begin{cases} 1 & \text{if } z \geq \theta \\ -1 & \text{otherwise} \end{cases}.$$

For simplicity, we can bring the threshold θ to the left side of the equation and define a weight-zero as $w_0 = -\theta$ and $x_0 = 1$, so that we write \mathbf{z} in a more compact form

$$z = w_0x_0 + w_1x_1 + \dots + w_mx_m = \mathbf{w}^T \mathbf{x}$$

and

$$\phi(z) = \begin{cases} 1 & \text{if } z \geq 0 \\ -1 & \text{otherwise} \end{cases}.$$

In the following sections, we will often make use of basic notations from linear algebra. For example, we will abbreviate the sum of the products of the values in \mathbf{x} and \mathbf{w} using a *vector dot product*, whereas superscript T stands for *transpose*, which is an operation that transforms a column vector into a row vector and vice versa:

$$z = w_0x_0 + w_1x_1 + \cdots + w_mx_m = \mathbf{w}^T \mathbf{x} = \sum_{j=0}^m \mathbf{w}_j \mathbf{x}_j = \mathbf{w}^T \mathbf{x}.$$

For example:

$$\begin{bmatrix} 1 & 2 & 3 \end{bmatrix} \times \begin{bmatrix} 4 \\ 5 \\ 6 \end{bmatrix} = 1 \times 4 + 2 \times 5 + 3 \times 6 = 32.$$

Furthermore, the transpose operation can also be applied to a matrix to reflect it over its diagonal, for example:

$$\begin{bmatrix} 1 & 2 \\ 3 & 4 \\ 5 & 6 \end{bmatrix}^T = \begin{bmatrix} 1 & 3 & 5 \\ 2 & 4 & 6 \end{bmatrix}$$

Rosenblatt’s initial perceptron rule is fairly simple and can be summarized by the following steps:

1. Initialize the weights to 0 or small random numbers.
2. For each training sample $\mathbf{x}^{(i)}$, perform the following steps:
 - (a) Compute the output value \hat{y} .
 - (b) Update the weights.

Here, the output value is the class label predicted by the unit step function that we defined earlier, and the simultaneous update of each weight w_j in the weight vector \mathbf{w} can be more formally written as:

$$w_j := w_j + \Delta w_j$$

The value of Δw_j , which is used to update the weight w_j , is calculated by the perceptron rule:

$$\Delta w_j = \eta \left(y^{(i)} - \hat{y}^{(i)} \right) x_j^{(i)}$$

Where η is the learning rate (a constant between 0.0 and 1.0), $y^{(i)}$ is the true class label of the i th training sample, and $\hat{y}^{(i)}$ is the predicted class label. It is important to note that all weights in the weight vector are being updated simultaneously, which means that we don't recompute $\hat{y}^{(i)}$ before all of the weights Δw_j were updated. Concretely, for a 2D dataset, we would write the update as follows:

$$\Delta w_0 = \eta \left(y^{(i)} - \hat{y}^{(i)} \right)$$

$$\Delta w_1 = \eta \left(y^{(i)} - \hat{y}^{(i)} \right) x_1^{(i)}$$

$$\Delta w_2 = \eta \left(y^{(i)} - \hat{y}^{(i)} \right) x_2^{(i)}$$

Before we implement the perceptron rule in Python, let us make a simple thought experiment to illustrate how beautifully simple this learning rule really is. In the two scenarios where the perceptron predicts the class label correctly, the weights remain unchanged:

$$\Delta w_j = \eta \left(-1 - -1 \right) x_j^{(i)} = 0$$

$$\Delta w_j = \eta \left(1 - 1 \right) x_j^{(i)} = 0$$

However, in the case of a wrong prediction, the weights are being pushed towards the direction of the positive or negative target class, respectively:

$$\Delta w_j = \eta \left(1 - -1 \right) x_j^{(i)} = \eta(2)x_j^{(i)}$$

$$\Delta w_j = \eta \left(-1 - 1 \right) x_j^{(i)} = \eta(-2)x_j^{(i)}$$

To get a better intuition for the multiplicative factor $x_j^{(i)}$, let us go through another simple example, where:

$$y^{(i)} = +1, \quad \hat{y}^{(i)} = -1, \quad \eta = 1$$

Let's assume that $x_j^{(i)} = 0.5$ and we misclassify this sample as -1 . In this case, we would increase the corresponding weight by 1 so that the activation $x_j^{(i)} \times w_j^{(i)}$ will be more positive the next time we encounter this sample and thus will be more likely to be above the threshold of the unit step function to classify the sample as $+1$:

$$\Delta w_j^{(i)} = (1 - -1)0.5 = (2)0.5 = 1$$

The weight update is proportional to the value of $x_j^{(i)}$. For example, if we have another sample $x_j^{(i)} = 2$ that is incorrectly classified as -1 , we'd push the decision boundary by an even larger extent to classify this sample correctly the next time:

$$\Delta w_j^{(i)} = (1 - (-1))2 = (2)2 = 4.$$

2.2 Implementing a perceptron learning algorithm in Python

2.2.1 Training a perceptron model on the Iris dataset

2.3 Adaptive linear neurons and the convergence of learning

The key difference between the Adaline rule (also known as the Widrow-Hoff rule) and Rosenblatt's perceptron is that the weights are updated based on a linear activation function rather than a unit step function like in the perceptron. In Adaline, this linear activation function ϕz is simply the identity function of the net input so that

$$\phi(\mathbf{w}^T \mathbf{x}) = \mathbf{w}^T \mathbf{x}$$

2.3.1 Minimizing cost functions with gradient descent

One of the key ingredients of supervised machine learning algorithms is to define an objective function that is to be optimized during the learning process. This objective function is often a cost function that we want to minimize. In the case of Adaline, we can define the cost function $J(\cdot)$ to learn the weights as the Sum of Squared Errors (SSE) between the calculated outcomes and the true class labels

$$J(\mathbf{w}) = \frac{1}{2} \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right)^2.$$

Using gradient descent, we can now update the weights by taking a step away from the gradient $\nabla J(\mathbf{w})$ of our cost function $J(\cdot)$:

$$\mathbf{w} := \mathbf{w} + \Delta \mathbf{w}.$$

To compute the gradient of the cost function, we need to compute the partial derivative of the cost function with respect to each weight w_j ,

$$\frac{\partial J}{\partial w_j} = - \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right) x_j^{(i)},$$

so that we can write the update of weight w_j as

$$\Delta w_j = -\eta \frac{\partial J}{\partial w_j} = \eta \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right) x_j^{(i)}$$

Since we update all weights simultaneously, our Adaline learning rule becomes

$$\mathbf{w} := \mathbf{w} + \Delta \mathbf{w}.$$

For those who are familiar with calculus, the partial derivative of the SSE cost function with respect to the j th weight in can be obtained as follows:

$$\begin{aligned} \frac{\partial J}{\partial w_j} &= \frac{\partial}{\partial w_j} \frac{1}{2} \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right)^2 \\ &= \frac{1}{2} \frac{\partial}{\partial w_j} \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right)^2 \\ &= \frac{1}{2} \sum_i 2(y^{(i)} - \phi(z^{(i)})) \frac{\partial}{\partial w_j} (y^{(i)} - \phi(z^{(i)})) \\ &= \sum_i (y^{(i)} - \phi(z^{(i)})) \frac{\partial}{\partial w_j} \left(y^{(i)} - \sum_i (w_j^{(i)} x_j^{(i)}) \right) \\ &= \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right) \left(-x_j^{(i)} \right) \\ &= -\sum_i \left(y^{(i)} - \phi(z^{(i)}) \right) x_j^{(i)} \end{aligned}$$

? Performing a matrix-vector multiplication is similar to calculating a vector dot product where each row in the matrix is treated as a single row vector. This vectorized approach represents a more compact notation and results in a more efficient computation using NumPy. For example:

$$\begin{bmatrix} 1 & 2 & 3 \\ 4 & 5 & 6 \end{bmatrix} \times \begin{bmatrix} 7 \\ 8 \\ 9 \end{bmatrix} = \begin{bmatrix} 1 \times 7 + 2 \times 8 + 3 \times 9 \\ 4 \times 7 + 5 \times 8 + 6 \times 9 \end{bmatrix} = \begin{bmatrix} 50 \\ 122 \end{bmatrix}$$

2.3.2 Implementing an Adaptive Linear Neuron in Python

Here, we will use a feature scaling method called standardization, which gives our data the property of a standard normal distribution. The mean of each feature is centered at value 0 and the feature column has a standard deviation of 1. For example, to standardize the j th feature, we simply need to subtract the sample mean μ_j from every training sample and divide it by its standard deviation σ_j :

$$\mathbf{x}'_j = \frac{\mathbf{x} - \mu_j}{\sigma_j}.$$

Here \mathbf{x}_j is a vector consisting of the j th feature values of all training samples n .

2.3.3 Large scale machine learning and stochastic gradient descent

A popular alternative to the batch gradient descent algorithm is stochastic gradient descent, sometimes also called iterative or on-line gradient descent. Instead of updating the weights based on the sum of the accumulated errors over all samples $\mathbf{x}^{(i)}$:

$$\Delta \mathbf{w} = \eta \sum_i \left(y^{(i)} - \phi(z^{(i)}) \right) \mathbf{x}^{(i)}.$$

We update the weights incrementally for each training sample:

$$\Delta \mathbf{w} = \eta \left(y^{(i)} - \phi(z^{(i)}) \right) \mathbf{x}^{(i)}.$$

2.4 Summary

Chapter 3

A Tour of Machine Learning Classifiers Using Scikit-learn

3.1 Choosing a classification algorithm

3.2 First steps with scikit-learn

3.2.1 Training a perceptron via scikit-learn

3.3 Modeling class probabilities via logistic regression

3.3.1 Logistic regression intuition and conditional probabilities

The odds ratio can be written as

$$\frac{p}{(1-p)},$$

where p stands for the probability of the positive (1? p) event. The term positive event does not necessarily mean good, but refers to the event that we want to predict, for example, the probability that a patient has a certain disease; we can think of the positive event as class label $y = 1$. We can then further define the logit function, which is simply the logarithm of the odds ratio (log-odds):

$$\text{logit}(p) = \log \frac{p}{1-p}$$

The logit function takes input values in the range 0 to 1 and transforms them to values over the entire real number range, which we can use to express a linear relationship between feature values and the log-odds:

$$\text{logit}(p(y = 1|\mathbf{x})) = w_0x_0 + w_1x_1 + \cdots + x_mw_m = \sum_{i=0}^m w_ix_i = \mathbf{w}^T \mathbf{x}.$$

Here, $p(y = 1|\mathbf{x})$ is the conditional probability that a particular sample belongs to class 1 given its features \mathbf{x} . Now what we are actually interested in is predicting the probability that a certain sample belongs to a particular class, which is the inverse form of the logit function. It is also called the logistic function, sometimes simply abbreviated as sigmoid function due to its characteristic S-shape

$$\phi(z) = \frac{1}{1 + e^{-z}}.$$

The output of the sigmoid function is then interpreted as the probability of particular sample belonging to class 1

$$\phi(z) = P(y = 1|\mathbf{x}; \mathbf{w})$$

given its features x parameterized by the weights w . For example, if we compute $\phi(z) = 0.8$ for a particular flower sample, it means that the chance that this sample is an Iris-Versicolor over is 80 percent. Similarly, the probability that this over is an Iris-Setosa over can be calculated as $P(y = 0|\mathbf{x}; \mathbf{w}) = 1 - P(y = 1|\mathbf{x}; \mathbf{w}) = 0.2$ or 20 percent. The predicted probability can then simply be converted into a binary outcome via a quantizer (unit step function):

$$\hat{y} = \begin{cases} 1 & \text{if } \phi(z) \geq 0.5 \\ 0 & \text{otherwise} \end{cases}.$$

If we look at the preceding sigmoid plot, this is equivalent to the following:

$$\hat{y} = \begin{cases} 1 & \text{if } \phi(z) \geq 0.0 \\ 0 & \text{otherwise} \end{cases}.$$

3.3.2 Learning the weights of the logistic cost function

In the previous chapter, we defined the sum-squared-error cost function:

$$J(\mathbf{w}) = \frac{1}{2} \sum_i \left(\phi(z^{(i)}) - y^{(i)} \right)^2.$$

We minimized this in order to learn the weights \mathbf{w} for our Adaline classification model. To explain how we can derive the cost function for logistic regression, let's first define the likelihood L that we want to maximize when we build a

logistic regression model, assuming that the individual samples in our dataset are independent of one another. The formula is as follows:

$$L(\mathbf{w}) = P(\mathbf{y}|\mathbf{x}; \mathbf{w}) = \prod_{i=1}^n P(y^{(i)}|x^{(i)}; \mathbf{w}) = \prod_{i=1}^n \left(\phi(z^{(i)}) \right)^{y^{(i)}} \left(1 - \phi(z^{(i)}) \right)^{1-y^{(i)}}$$

In practice, it is easier to maximize the (natural) log of this equation, which is called the log-likelihood function:

$$l(\mathbf{w}) = \log L(\mathbf{w}) = \sum_{i=1}^n \left[y^{(i)} \log \left(\phi(z^{(i)}) \right) + \left(1 - y^{(i)} \right) \log \left(1 - \phi(z^{(i)}) \right) \right]$$

Firstly, applying the log function reduces the potential for numerical underflow, which can occur if the likelihoods are very small. Secondly, we can convert the product of factors into a summation of factors, which makes it easier to obtain the derivative of this function via the addition trick, as you may remember from calculus.

Now we could use an optimization algorithm such as gradient ascent to maximize this log-likelihood function. Alternatively, let's rewrite the log-likelihood as a cost function $J(\cdot)$ that can be minimized using gradient descent as in *Chapter 2, Training Machine Learning Algorithms for Classification*:

$$J(\mathbf{w}) = \sum_{i=1}^n \left[-y^{(i)} \log \left(\phi(z^{(i)}) \right) - \left(1 - y^{(i)} \right) \log \left(1 - \phi(z^{(i)}) \right) \right]$$

To get a better grasp on this cost function, let's take a look at the cost that we calculate for one single-sample instance:

$$J(\phi(z), y; \mathbf{w}) = -y \log(\phi(z)) - (1 - y) \log(1 - \phi(z)).$$

Looking at the preceding equation, we can see that the first term becomes zero if $y = 0$, and the second term becomes zero if $y = 1$, respectively:

$$J(\phi(z), y; \mathbf{w}) = \begin{cases} -\log(\phi(z)) & \text{if } y = 1 \\ -\log(1 - \phi(z)) & \text{if } y = 0 \end{cases}$$

3.3.3 Training a logistic regression model with scikit-learn

If we were to implement logistic regression ourselves, we could simply substitute the cost function $J(\cdot)$ in our Adaline implementation from *Chapter 2, Training Machine Learning Algorithms for Classification*, by the new cost function:

$$J(\mathbf{w}) = \sum_{i=1}^n \left[-y^{(i)} \log \left(\phi(z^{(i)}) \right) - \left(1 - y^{(i)} \right) \log \left(1 - \phi(z^{(i)}) \right) \right]$$

We can show that the weight update in logistic regression via gradient descent is indeed equal to the equation that we used in Adaline in *Chapter 2, Training Machine Learning Algorithms for Classification*. Let's start by calculating the partial derivative of the log-likelihood function with respect to the j th weight:

$$\frac{\partial}{\partial w_j} l(\mathbf{w}) = \left(y \frac{1}{\phi(z)} - (1-y) \frac{1}{1-\phi(z)} \right) \frac{\partial}{\partial w_j} \phi(z)$$

Before we continue, let's calculate the partial derivative of the sigmoid function first:

$$\begin{aligned} \frac{\partial}{\partial z} \phi(z) &= \frac{\partial}{\partial z} \frac{1}{1+e^{-z}} = \frac{1}{(1+e^{-z})^2} e^{-z} = \frac{1}{1+e^{-z}} = \frac{1}{1+e^{-z}} \left(1 - \frac{1}{1+e^{-z}} \right) \\ &= \phi(z)(1-\phi(z)). \end{aligned}$$

Now we can resubstitute $\frac{\partial}{\partial z} \phi(z) = \phi(z)(1-\phi(z))$ in our first equation to obtain the following:

$$\begin{aligned} &\left(y \frac{1}{\phi(z)} - (1-y) \frac{1}{1-\phi(z)} \right) \frac{\partial}{\partial w_j} \phi(z) \\ &= \left(y \frac{1}{\phi(z)} - (1-y) \frac{1}{1-\phi(z)} \right) \phi(z)(1-\phi(z)) \frac{\partial}{\partial w_j} z \\ &= \left(y(1-\phi(z)) - (1-y)\phi(z) \right) x_j \\ &= (y - \phi(z)) x_j \end{aligned}$$

Remember that the goal is to find the weights that maximize the log-likelihood so that we would perform the update for each weight as follows:

$$w_j := w_j + \eta \sum_{i=1}^n \left(y^{(i)} - \phi(z^{(i)}) \right) x_j^{(i)}$$

Since we update all weights simultaneously, we can write the general update rule as follows:

$$\mathbf{w} := \mathbf{w} + \Delta \mathbf{w}$$

We define $\Delta \mathbf{w}$ as follows:

$$\Delta \mathbf{w} = \eta \nabla l(\mathbf{w})$$

Since maximizing the log-likelihood is equal to minimizing the cost function $J(\cdot)$ that we defined earlier, we can write the gradient descent update rule as follows:

$$\Delta w_j = -\eta \frac{\partial J}{\partial w_j} = \eta \sum_{i=1}^n \left(y^{(i)} - \phi(z^{(i)}) \right) x_j^{(i)}$$

$$\mathbf{w} := \mathbf{w} + \Delta \mathbf{w}, \Delta \mathbf{w} = -\eta \nabla J(\mathbf{w})$$

This is equal to the gradient descent rule in Adaline in *Chapter 2, Training Machine Learning Algorithms for Classification*.

3.3.4 Tackling overfitting via regularization

The most common form of regularization is the so-called L2 regularization (sometimes also called L2 shrinkage or weight decay), which can be written as follows:

$$\frac{\lambda}{2} \|\mathbf{w}\|^2 = \frac{\lambda}{2} \sum_{j=1}^m w_j^2$$

Here, λ is the so-called regularization parameter.

In order to apply regularization, we just need to add the regularization term to the cost function that we defined for logistic regression to shrink the weights:

$$J(\mathbf{w}) = \sum_{i=1}^n \left[-y^{(i)} \log(\phi(z^{(i)})) - (1 - y^{(i)}) \log(1 - \phi(z^{(i)})) \right] + \frac{\lambda}{2} \|\mathbf{w}\|^2$$

Via the regularization parameter λ , we can then control how well we fit the training data while keeping the weights small. By increasing the value of λ , we increase the regularization strength.

The parameter C that is implemented for the *LogisticRegression* class in scikit-learn comes from a convention in support vector machines, which will be the topic of the next section. C is directly related to the regularization parameter λ , which is its inverse:

$$C = \frac{1}{\lambda}$$

So, we can rewrite the regularized cost function of logistic regression as follows:

$$J(\mathbf{w}) = C \left[\sum_{i=1}^n \left(-y^{(i)} \log(\phi(z^{(i)})) - (1 - y^{(i)}) \log(1 - \phi(z^{(i)})) \right) \right] + \frac{1}{2} \|\mathbf{w}\|^2$$

3.4 Maximum margin classification with support vector machines

3.4.1 Maximum margin intuition

To get an intuition for the margin maximization, let's take a closer look at those *positive* and *negative* hyperplanes that are parallel to the decision boundary, which can be expressed as follows:

$$w_0 + \mathbf{w}^T \mathbf{x}_{pos} = 1 \quad (1)$$

$$w_0 + \mathbf{w}^T \mathbf{x}_{neg} = -1 \quad (2)$$

If we subtract those two linear equations (1) and (2) from each other, we get:

$$\Rightarrow \mathbf{w}^T (\mathbf{x}_{pos} - \mathbf{x}_{neg}) = 2$$

We can normalize this by the length of the vector \mathbf{w} , which is defined as follows:

$$\|\mathbf{w}\| = \sqrt{\sum_{j=1}^m w_j^2}$$

So we arrive at the following equation:

$$\frac{\mathbf{w}^T (\mathbf{x}_{pos} - \mathbf{x}_{neg})}{\|\mathbf{w}\|} = \frac{2}{\|\mathbf{w}\|}$$

The left side of the preceding equation can then be interpreted as the distance between the positive and negative hyperplane, which is the so-called margin that we want to maximize.

Now the objective function of the SVM becomes the maximization of this margin by maximizing $\frac{2}{\|\mathbf{w}\|}$ under the constraint that the samples are classified correctly, which can be written as follows:

$$w_0 + \mathbf{w}^T \mathbf{x}^{(i)} \geq 1 \text{ if } y^{(i)} = 1$$

$$w_0 + \mathbf{w}^T \mathbf{x}^{(i)} < -1 \text{ if } y^{(i)} = -1$$

These two equations basically say that all negative samples should fall on one side of the negative hyperplane, whereas all the positive samples should fall behind the positive hyperplane. This can also be written more compactly as follows:

$$y^{(i)} (w_0 + \mathbf{w}^T \mathbf{x}^{(i)}) \geq 1 \quad \forall_i$$

In practice, though, it is easier to minimize the reciprocal term $\frac{1}{2} \|\mathbf{w}\|^2$, which can be solved by quadratic programming.

3.4.2 Dealing with the nonlinearly separable case using slack variables

The motivation for introducing the slack variable ξ was that the linear constraints need to be relaxed for nonlinearly separable data to allow convergence of the optimization in the presence of misclassifications under the appropriate cost penalization. The positive-values slack variable is simply added to the linear constraints:

$$\mathbf{w}^T \mathbf{x}^{(i)} \geq 1 - \xi^{(i)} \text{ if } y^{(i)} = 1$$

$$\mathbf{w}^T \mathbf{x}^{(i)} < -1 + \xi^{(i)} \text{ if } y^{(i)} = -1$$

So the new objective to be minimized (subject to the preceding constraints) becomes:

$$\frac{1}{2} \|\mathbf{w}\|^2 + C \left(\sum_i \xi^{(i)} \right)$$

3.4.3 Alternative implementations in scikit-learn

3.5 Solving nonlinear problems using a kernel SVM

As shown in the next figure, we can transform a two-dimensional dataset onto a new three-dimensional feature space where the classes become separable via the following projection:

$$\phi(x_1, x_2) = (z_1, z_2, z_3) = (x_1, x_2, x_1^2 + x_2^2)$$

3.5.1 Using the kernel trick to find separating hyperplanes in higher dimensional space

To solve a nonlinear problem using an SVM, we transform the training data onto a higher dimensional feature space via a mapping function $\phi(\cdot)$ and train a linear SVM model to classify the data in this new feature space. Then we can use the same mapping function $\phi(\cdot)$ to transform new, unseen data to classify it using the linear SVM model.

However, one problem with this mapping approach is that the construction of the new features is computationally very expensive, especially if we are dealing with high-dimensional data. This is where the so-called kernel trick comes into play. Although we didn't go into much detail about how to solve the quadratic programming task to train an SVM, in practice all we need is to replace the dot product

$$\mathbf{x}^{(i)T} \mathbf{x}^{(j)} \text{ by } \phi(\mathbf{x}^{(i)})^T \phi(\mathbf{x}^{(j)})$$

In order to save the expensive step of calculating this dot product between two points explicitly, we define a so-called kernel function:

$$k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \phi(\mathbf{x}^{(i)})^T \phi(\mathbf{x}^{(j)})$$

One of the most widely used kernels is the *Radial Basis Function kernel* (RBF kernel) or Gaussian kernel:

$$k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp\left(-\frac{\|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|^2}{2\sigma^2}\right)$$

This is often simplified to:

$$k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp\left(-\gamma \|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|^2\right)$$

Here, $\gamma = \frac{1}{2\sigma^2}$ is a free parameter that is to be optimized.

3.6 Decision tree learning

In order to split the nodes at the most informative features, we need to define an objective function that we want to optimize via the tree learning algorithm. Here, our objective function is to maximize the information gain at each split, which we define as follows:

$$IG(D_p, f) = I(D_p) - \sum_{j=1}^m \frac{N_j}{N_p} I(D_j)$$

Here, f is the feature to perform the split, D_p and D_j are the dataset of the parent p and j th child node; I is our impurity measure, N_p is the total number of samples at the parent node, and N_j is the number of samples at the j th child node. As we can see, the information gain is simply the difference between the impurity of the parent node and the sum of the child node impurities; the lower the impurity of the child nodes, the larger the information gain. However, for simplicity and to reduce the combinatorial search space, most libraries (including scikit-learn) implement binary decision trees. This means that each parent node is split into two child nodes, D_{left} and D_{right} :

$$IG(D_p, f) = I(D_p) - \frac{N_{left}}{N_p} I(D_{left}) - \frac{N_{right}}{N_p} I(D_{right})$$

Now, the three impurity measures or splitting criteria that are commonly used in binary decision trees are *Gini impurity* (I_G), *Entropy* (I_H) and the *classification error* (I_E). Let's start with the definition of Entropy for all non-empty classes $p(i|t) \neq 0$:

$$I_H(t) = -\sum_{i=1}^c p(i|t) \log_2 p(i|t)$$

Here, $p(i|t)$ is the proportion of the samples that belongs to class i for a particular node t . The entropy is therefore 0 if all samples at a node belong to the same class, and the entropy is maximal if we have a uniform class distribution. For example, in a binary class setting, the entropy is 0 if $p(i = 1|t) = 1$ or $p(i = 0|t) = 0$. If the classes are distributed uniformly with $p(i = 1|t) = 0.5$ and $p(i = 0|t) = 0.5$, the entropy is 1. Therefore, we can say that the entropy criterion attempts to maximize the mutual information in the tree. Intuitively, the Gini impurity can be understood as a criterion to minimize the probability of misclassification:

$$I_G(t) = \sum_{i=1}^c p(i|t)(1 - p(i|t)) = 1 - \sum_{i=1}^c p(i|t)^2$$

Similar to entropy, the Gini impurity is maximal if the classes are perfectly mixed, for example, in a binary class setting ($c = 2$):

$$1 - \sum_{i=1}^c 0.5^2 = 0.5.$$

...

Another impurity measure is the classification error:

$$I_E = 1 - \max\{p(i|t)\}$$

3.6.1 Maximizing information gain – getting the most bang for the buck

3.6.2 Building a decision tree

3.6.3 Combining weak to strong learners via random forests

3.7 K-nearest neighbors – a lazy learning algorithm

The *minkowski* distance that we used in the previous code example is just a generalization of the Euclidean and Manhattan distances that can be written as follows:

$$d(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \sqrt[p]{\sum_k |x_k^{(i)} - x_k^{(j)}|^p}$$

3.8 Summary

Chapter 4

Building Good Training Sets – Data Pre-Processing

4.1 Dealing with missing data

4.1.1 Eliminating samples or features with missing values

4.1.2 Imputing missing values

4.1.3 Understanding the scikit-learn estimator API

4.2 Handling categorical data

4.2.1 Mapping ordinal features

4.2.2 Encoding class labels

4.2.3 Performing one-hot encoding on nominal features

4.3 Partitioning a dataset in training and test sets

4.4 Bringing features onto the same scale

Now, there are two common approaches to bringing different features onto the same scale: *normalization* and *standardization*. Those terms are often used quite loosely in different fields, and the meaning has to be derived from the context. Most often, *normalization* refers to the rescaling of the features to a range of $[0, 1]$, which is a special case of min-max scaling. To normalize our data, we can simply apply the min-max scaling to each feature column, where the new value $x_{norm}^{(i)}$ of a sample $x^{(i)}$:

$$x_{norm}^{(i)} = \frac{x^{(i)} - \mathbf{x}_{min}}{\mathbf{x}_{max} - \mathbf{x}_{min}}$$

Here, $x^{(i)}$ is a particular sample, x_{min} is the smallest value in a feature column, and x_{max} the largest value, respectively.

[...] Furthermore, standardization maintains useful information about outliers and makes the algorithm less sensitive to them in contrast to min-max scaling, which scales the data to a limited range of values.

The procedure of standardization can be expressed by the following equation:

$$x_{std}^{(i)} = \frac{x^{(i)} - \mu_x}{\sigma_x}$$

Here, μ_x is the sample mean of a particular feature column and σ_x the corresponding standard deviation, respectively.

4.5 Selecting meaningful features

4.5.1 Sparse solutions with L1 regularization

We recall from *Chapter 3, A Tour of Machine Learning Classifiers Using Scikit-learn*, that L2 regularization is one approach to reduce the complexity of a model by penalizing large individual weights, where we defined the L2 norm of our weight vector \mathbf{w} as follows:

$$L2 : \|\mathbf{w}\|_2^2 = \sum_{j=1}^m w_j^2$$

Another approach to reduce the model complexity is the related *L1 regularization*:

$$L1 : \|\mathbf{w}\|_1 = \sum_{j=1}^m |w_j|$$

4.5.2 Sequential feature selection algorithms

Based on the preceding definition of SBS, we can outline the algorithm in 4 simple steps:

1. Initialize the algorithm with $k = d$, where d is the dimensionality of the full feature space \mathbf{X}_d
2. Determine the feature x^- that maximizes the criterion $x^- = \arg \max J(\mathbf{X}_k - x)$, where $x \in \mathbf{X}_k$.
3. Remove the feature x^- from the feature set: $\mathbf{X}_{k-1} := \mathbf{X}_k - x^-$; $k := k-1$.
4. Terminate if k equals the number of desired features, if not, got to step 2.

4.6 Assessing feature importance with random forests

4.7 Summary

Chapter 5

Compressing Data via Dimensionality Reduction

5.1 Unsupervised dimensionality reduction via principal component analysis

When we use PCA for dimensionality reduction, we construct a $d \times k$ -dimensional transformation matrix \mathbf{W} that allows us to map a sample vector \mathbf{x} onto a new k -dimensional feature subspace that has fewer dimensions than the original d -dimensional feature space:

$$\mathbf{x} = [x_1, x_2, \dots, x_d], \mathbf{x} \in \mathbb{R}^d$$

$$\downarrow \mathbf{x}\mathbf{W}, \quad \mathbf{W} \in \mathbb{R}^{d \times k}$$

$$\mathbf{z} = [z_1, z_2, \dots, z_k], \quad \mathbf{z} \in \mathbb{R}^k$$

As a result of transforming the original d -dimensional data onto this new k -dimensional subspace (typically $k \ll d$), the 1st principal component will have the largest possible variance, and all consequent principal components will have the largest possible variance given that they are uncorrelated (orthogonal) to the other principal components. Note that the PCA directions are highly sensitive to data scaling, and we need to standardize the features prior to PCA if the features were measured on different scales and we want to assign equal importance to all features.

Before looking at the PCA algorithm for dimensionality reduction in more detail, let's summarize the approach in a few simple steps:

1. Standardize the d -dimensional dataset.
2. Construct the covariance matrix.

3. Decompose the covariance matrix into its eigenvectors and eigenvalues.
4. Select k eigenvectors that correspond to the k largest eigenvalues, where k is the dimensionality of the new feature subspace ($k \leq d$).
5. Construct a projection matrix \mathbf{W} from the "top" k eigenvectors.
6. Transform the d -dimensional input dataset \mathbf{X} using the projection matrix \mathbf{W} to obtain the new k -dimensional feature subspace.

5.1.1 Total and explained variance

After completing the mandatory preprocessing steps by executing the preceding code, let's advance to the second step: constructing the covariance matrix. The symmetric $d \times d$ -dimensional covariance matrix, where d is the number of dimensions in the dataset, stores the pairwise covariances between the different features. For example, the covariance between two features \mathbf{x}_j and \mathbf{x}_k on the population level can be calculated via the following equation:

$$\sigma_{jk} = \frac{1}{n} \sum_{i=1}^n (x_j^{(i)} - \mu_j)(x_k^{(i)} - \mu_k)$$

Here, μ_j and μ_k are the sample means of feature j and k , respectively. [...] For example, a covariance matrix of three features can then be written as

$$\Sigma = \begin{bmatrix} \sigma_1^2 & \sigma_{12} & \sigma_{13} \\ \sigma_{21} & \sigma_2^2 & \sigma_{23} \\ \sigma_{31} & \sigma_{32} & \sigma_3^2 \end{bmatrix}$$

[...] an eigenvector \mathbf{v} satisfies the following condition:

$$\Sigma \mathbf{v} = \lambda \mathbf{v}$$

Here, λ is a scalar: the eigenvalue.

...

The variance explained ratio of an eigenvalue λ_j is simply the fraction of an eigenvalue λ_j and the total sum of the eigenvalues:

$$\frac{\lambda_j}{\sum_{j=1}^d \lambda_j}$$

5.1.2 Feature transformation

Using the projection matrix, we can now transform a sample \mathbf{x} onto the PCA subspace obtaining \mathbf{x}' , a now two-dimensional sample vector consisting of two new features:

$$\mathbf{x}' = \mathbf{x}\mathbf{W}$$

5.1.3 Principal component analysis in scikit-learn

5.2 Supervised data compression via linear discriminant analysis

Before we take a look into the inner workings of LDA in the following subsections, let's summarize the key steps of the LDA approach:

1. Standardize the d -dimensional dataset (d is the number of features).
2. For each class, compute the d dimensional mean vector.
3. Construct the between-class scatter matrix \mathbf{S}_B and the within-class scatter matrix \mathbf{S}_W .
4. Compute the eigenvectors and corresponding eigenvalues of the matrix $\mathbf{S}_W^{-1}\mathbf{S}_B$.
5. Choose the k eigenvectors that correspond to the k largest eigenvalues to construct a $d \times k$ -dimensional transformation matrix \mathbf{W} ; the eigenvectors are the columns of this matrix.
6. Project the samples onto the new feature subspace using the transformation matrix \mathbf{W} .

5.2.1 Computing the scatter matrices

Each mean vector \mathbf{m}_i stores the mean feature value μ_m with respect to the samples of class i :

$$\mathbf{m}_i = \frac{1}{n_i} \sum_{x \in D_i}^c \mathbf{x}_m$$

This results in three mean vectors:

$$\mathbf{m}_i = \begin{bmatrix} \mu_{i,\text{alcohol}} \\ \mu_{i,\text{malic-acid}} \\ \mu_{i,\text{proline}} \end{bmatrix}, i \in \{1, 2, 3\}$$

Using the mean vectors, we can now compute the within-class scatter matrix \mathbf{S}_W

$$\mathbf{S}_W = \sum_{i=1}^c \mathbf{S}_i$$

This is calculated by summing up the individual scatter matrices \mathbf{S}_i of each individual class i :

$$\mathbf{S}_i = \sum_{\mathbf{x} \in D_i}^c (\mathbf{x} - \mathbf{m}_i)(\mathbf{x} - \mathbf{m}_i)^T$$

The assumption that we are making when we are computing the scatter matrices is that the class labels in the training set are uniformly distributed. [...] Thus, we want to scale the individual scatter matrices \mathbf{S}_i before we sum them up as scatter matrix \mathbf{S}_W . When we divide the scatter matrices by the number of class samples N_i , we can see that computing the scatter matrix is in fact the same as computing the covariance matrix \mathbf{Sigma}_i . The covariance matrix is a normalized version of the scatter matrix:

$$\Sigma_i = \frac{1}{N_i} \mathbf{S}_i = \frac{1}{N_i} \sum_{\mathbf{x} \in D_i}^c (\mathbf{x} - \mathbf{m}_i)(\mathbf{x} - \mathbf{m}_i)^T$$

After we have computed the scaled within-class scatter matrix (or covariance matrix), we can move on to the next step and compute the between-class scatter matrix \mathbf{S}_B :

$$\mathbf{S}_B = \sum_{i=1}^c N_i (\mathbf{m}_i - \mathbf{m})(\mathbf{m}_i - \mathbf{m})^T$$

Here, \mathbf{m} is the overall mean that is computed, including samples from all classes.

5.2.2 Selecting linear discriminants for the new feature subspace

5.2.3 Projecting samples onto the new feature space

$$\mathbf{X}' = \mathbf{X}\mathbf{W}$$

5.2.4 LDA via scikit-learn

5.3 Using kernel principal component analysis for nonlinear mappings

5.3.1 Kernel functions and the kernel trick

To transform the samples $\mathbf{x} \in \mathbb{R}^d$ onto this higher k -dimensional subspace, we defined a nonlinear mapping function ϕ :

$$\phi: \mathbb{R}^d \rightarrow \mathbb{R}^k \quad (k \gg d)$$

We can think of ϕ as a function that creates nonlinear combinations of the original features to map the original d -dimensional dataset onto a larger, k -dimensional feature space. For example, if we had feature vector $\mathbf{x} \in \mathbb{R}^d$ (\mathbf{x} is a

column vector consisting of d features) with two dimensions ($d = 2$), a potential mapping onto a 3D space could be as follows:

$$\begin{aligned}\mathbf{x} &= [x_1, x_2]^T \\ &\downarrow \phi \\ \mathbf{z} &= \left[x_1^2, \sqrt{2x_1x_2}, x_2^2 \right]^T\end{aligned}$$

[...] We computed the covariance between two features k and j as follows:

$$\sigma_{jk} = \frac{1}{n} \sum_{i=1}^n (x_j^{(i)} - \mu_j)(x_k^{(i)} - \mu_k)$$

Since the standardizing of features centers them at mean zero, for instance, $\mu_j = 0$ and $\mu_k = 0$, we can simplify this equation as follows:

$$\sigma_{jk} = \frac{1}{n} \sum_{i=1}^n x_j^{(i)} x_k^{(i)}$$

Note that the preceding equation refers to the covariance between two features; now, let's write the general equation to calculate the covariance matrix Σ :

$$\Sigma = \frac{1}{n} \sum_{i=1}^n \mathbf{x}^{(i)} \mathbf{x}^{(i)T}$$

Bernhard Scholkopf generalized this approach (B. Scholkopf, A. Smola, and K.-R. Muller. *Kernel Principal Component Analysis*. pages 583-588, 1997) so that we can replace the dot products between samples in the original feature space by the nonlinear feature combinations via ϕ :

$$\Sigma = \frac{1}{n} \sum_{i=1}^n \phi(\mathbf{x}^{(i)}) \phi(\mathbf{x}^{(i)})^T$$

To obtain the eigenvectors?the principal components?from this covariance matrix, we have to solve the following equation:

$$\begin{aligned}\Sigma \mathbf{v} &= \lambda \mathbf{v} \\ \Rightarrow \frac{1}{n} \sum_{i=1}^n \phi(\mathbf{x}^{(i)}) \phi(\mathbf{x}^{(i)})^T \mathbf{v} &= \lambda \mathbf{v} \\ \Rightarrow \frac{1}{n\lambda} \sum_{i=1}^n \phi(\mathbf{x}^{(i)}) (\mathbf{x}^{(i)})^T \mathbf{v} &= \frac{1}{n} \sum_{i=1}^n \mathbf{a}^{(i)} \phi(\mathbf{x}^{(i)})\end{aligned}$$

Here, λ and \mathbf{v} are the eigenvalues and eigenvectors of the covariance matrix Σ , and \mathbf{a} can be obtained by extracting the eigenvectors of the kernel (similarity) matrix \mathbf{K} as we will see in the following paragraphs.

The derivation of the kernel matrix is as follows:

5.3.2 Implementing a kernel principal component analysis in Python

First, let's write the covariance matrix as in matrix notation, where $\phi(X)$ is an $n \times k$ -dimensional matrix:

$$\Sigma = \frac{1}{n} \sum_{i=1}^n \phi(\mathbf{x}^{(i)}) (\mathbf{x}^{(i)})^T = \frac{1}{n} \phi(\mathbf{X})^T \phi(\mathbf{X})$$

Now, we can write the eigenvector equation as follows:

$$\Sigma \mathbf{v} = \frac{1}{n} \sum_{i=1}^n \mathbf{a}^{(i)} \phi(\mathbf{x}^{(i)}) = \lambda \phi(\mathbf{X})^T \mathbf{a}$$

Since $\Sigma \mathbf{v} = \lambda \mathbf{v}$, we get:

$$\frac{1}{n} \phi(\mathbf{X})^T \phi(\mathbf{X}) \phi(\mathbf{X})^T \mathbf{a} = \lambda \phi(\mathbf{X})^T \mathbf{a}$$

Multiplying it by $\phi(\mathbf{X})$ on both sides yields the following result:

$$\begin{aligned} \frac{1}{n} \phi(\mathbf{X}) \phi(\mathbf{X})^T \phi(\mathbf{X}) \phi(\mathbf{X})^T \mathbf{a} &= \lambda \phi(\mathbf{X}) \phi(\mathbf{X})^T \mathbf{a} \\ \Rightarrow \frac{1}{n} \phi(\mathbf{X}) \phi(\mathbf{X})^T \mathbf{a} &= \lambda \mathbf{a} \\ \Rightarrow \frac{1}{n} \mathbf{K} \mathbf{a} &= \lambda \mathbf{a} \end{aligned}$$

Here, \mathbf{K} is the similarity (kernel) matrix:

$$\mathbf{K} = \phi(\mathbf{X}) \phi(\mathbf{X})^T$$

As we recall from the SVM section in *Chapter 3, A Tour of Machine Learning Classifiers Using Scikit-learn*, we use the kernel trick to avoid calculating the pairwise dot products of the samples \mathbf{x} under ϕ explicitly by using a kernel function $\kappa(\cdot)$ so that we don't need to calculate the eigenvectors explicitly:

[...] The most commonly used kernels are the following ones:

- The polynomial kernel:

$$\kappa(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = (\mathbf{x}^{(i)T} \mathbf{x}^{(j)} + \theta)^p$$

Here, θ is the threshold and p is the power that has to be specified by the user.

- the hyperbolic tangent (sigmoid) kernel:

$$\kappa(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \tanh(\eta \mathbf{x}^{(i)T} \mathbf{x}^{(j)} + \theta)$$

- The *Radial Basis Function (RBF)* or Gaussian kernel that we will use in the following examples in the next subsection:

$$\kappa(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp\left(-\frac{\|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|^2}{2\sigma^2}\right),$$

which is also often written as

$$\kappa(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp(-\gamma \|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|^2),$$

where $\gamma = \frac{1}{2\sigma^2}$.

To summarize what we have discussed so far, we can define the following three steps to implement an RBF kernel PCA:

1. We compute the kernel (similarity) matrix \mathbf{K} , where we need to calculate the following:

$$\kappa(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp(-\gamma \|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|^2)$$

We do this for each pair of samples:

$$\mathbf{K} = \begin{bmatrix} \kappa(\mathbf{x}^{(1)}, \mathbf{x}^{(1)}) & \kappa(\mathbf{x}^{(1)}, \mathbf{x}^{(2)}) & \dots & \kappa(\mathbf{x}^{(1)}, \mathbf{x}^{(n)}) \\ \kappa(\mathbf{x}^{(2)}, \mathbf{x}^{(1)}) & \kappa(\mathbf{x}^{(2)}, \mathbf{x}^{(2)}) & \dots & \kappa(\mathbf{x}^{(2)}, \mathbf{x}^{(n)}) \\ \vdots & \vdots & \ddots & \vdots \\ \kappa(\mathbf{x}^{(n)}, \mathbf{x}^{(1)}) & \kappa(\mathbf{x}^{(n)}, \mathbf{x}^{(2)}) & \dots & \kappa(\mathbf{x}^{(n)}, \mathbf{x}^{(n)}) \end{bmatrix}.$$

For example, if our dataset contains 100 training samples, the symmetric kernel matrix of the pair-wise similarities would be 100×100 dimensional.

2. We center the kernel matrix \mathbf{K} using the following equation:

$$\mathbf{K}' = \mathbf{K} - \mathbf{1}_n \mathbf{K} - \mathbf{K} \mathbf{1}_n + \mathbf{1}_n \mathbf{K} \mathbf{1}_n$$

Here, $\mathbf{1}_n$ is an $n \times n$ -dimensional matrix (the same dimensions as the kernel matrix) where all values are equal to $\frac{1}{n}$.

3. We collect the top k eigenvectors of the centered kernel matrix based on their corresponding eigenvalues, which are ranked by decreasing magnitude. In contrast to standard PCA, the eigenvectors are not the principal component axes but the samples projected onto those axes

Example 1 – separating half-moon shapes

Example 2 – separating concentric circles

5.3.3 Projecting new data points

[...] Thus, if we want to project a new sample \mathbf{x}' onto this principal component axis, we'd need to compute the following:

$$\phi(\mathbf{x}')^T \mathbf{v}$$

Fortunately, we can use the kernel trick so that we don't have to calculate the projection $\phi(\mathbf{x}')^T \mathbf{v}$ explicitly. However, it is worth noting that kernel PCA, in contrast to standard PCA, is a memory-based method, which means that we have to reuse the original training set each time to project new samples. We have to calculate the pairwise RBF kernel (similarity) between each i th sample in the training dataset and the new sample \mathbf{x}' :

$$\begin{aligned} \phi(\mathbf{x}')^T \mathbf{v} &= \sum_i \mathbf{a}^{(i)} \phi(\mathbf{x}')^T \phi(\mathbf{x}^{(i)}) \\ &= \sum_i \mathbf{a}^{(i)} k(\mathbf{x}', \mathbf{x}^{(i)})^T \end{aligned}$$

Here, eigenvectors \mathbf{a} and eigenvalues λ of the Kernel matrix \mathbf{K} satisfy the following condition in the equation

$$\mathbf{K}\mathbf{a} = \lambda\mathbf{a}$$

5.3.4 Kernel principal component analysis in scikit-learn

5.4 Summary

Chapter 6

Learning Best Practices for Model Evaluation and Hyperparameter Tuning

6.1 Streamlining workflows with pipelines

6.1.1 Loading the Breast Cancer Wisconsin dataset

6.1.2 Combining transformers and estimators in a pipeline

6.2 Using k-fold cross-validation to assess model performance

6.2.1 The holdout method

6.2.2 K-fold cross-validation

6.3 Debugging algorithms with learning and validation curves

6.3.1 Diagnosing bias and variance problems with learning curves

6.3.2 Addressing overfitting and underfitting with validation curves

6.4 Fine-tuning machine learning models via grid search

6.4.1 Tuning hyperparameters via grid search

6.4.2 Algorithm selection with nested cross-validation³²

6.5 Looking at different performance evaluation metrics

6.5.1 Reading a confusion matrix

6.5.2 Optimizing the precision and recall of a classification

the sum of all false predictions divided by the number of total predictions, and the accuracy is calculated as the sum of correct predictions divided by the total number of predictions, respectively:

$$ERR = \frac{FP + FN}{FP + FN + TP + TN}$$

(TP = true positives, FP = false positives, TN = true negatives, FN = false negatives)

The prediction accuracy can then be calculated directly from the error:

$$ACC = \frac{TP + TN}{FP + FN + TP + TN} = 1 - ERR$$

The true *positive rate* (TPR) and *false positive rate* (FPR) are performance metrics that are especially useful for imbalanced class problems:

$$FPR = \frac{FP}{N} = \frac{FP}{FP + TN}$$

$$TPR = \frac{TP}{P} = \frac{TP}{FN + TP}$$

Precision (PRE) and *recall* (REC) are performance metrics that are related to those true positive and true negative rates, and in fact, recall is synonymous to the true positive rate:

$$PRE = \frac{TP}{TP + FP}$$

$$REC = TPR = \frac{TP}{P} = \frac{TP}{FN + TP}$$

In practice, often a combination of precision and recall is used, the so-called *F1-score*:

$$F1 = 2 \times \frac{PRE \times REC}{PRE + REC}$$

6.5.3 Plotting a receiver operating characteristic

6.5.4 The scoring metrics for multiclass classification

the micro-average is calculated from the individual true positives, true negatives, false positives, and false negatives of the system. For example, the micro-average of the precision score in a k-class system can be calculated as follows:

$$PRE_{micro} = \frac{TP_1 + \dots + TP_k}{TP_1 + \dots + TP_k + FP_1 + \dots + FP_k}$$

The macro-average is simply calculated as the average scores of the different systems:

$$PRE_{macro} = \frac{PRE_1 + \dots + PRE_k}{k}$$

6.6 Summary

Chapter 7

Combining Different Models for Ensemble Learning

7.1 Learning with ensembles

To predict a class label via a simple majority or plurality voting, we combine the predicted class labels of each individual classifier C_j and select the class label \hat{y} that received the most votes:

$$\hat{y} = \text{mode}\{C_1(\mathbf{x}), C_2(\mathbf{x}), \dots, C_m(\mathbf{x})\}$$

For example, in a binary classification task where $class1 = -1$ and $class2 = +1$, we can write the majority vote prediction as follows:

$$C(\mathbf{x}) = \text{sign}\left[\sum_j^m C_j(\mathbf{x})\right] = \begin{cases} 1 & \text{if } \sum_j C_j(\mathbf{x}) \geq 0 \\ -1 & \text{otherwise} . \end{cases}$$

To illustrate why ensemble methods can work better than individual classifiers alone, let's apply the simple concepts of combinatorics. For the following example, we make the assumption that all n base classifiers for a binary classification task have an equal error rate ϵ . Furthermore, we assume that the classifiers are independent and the error rates are not correlated. Under those assumptions, we can simply express the error probability of an ensemble of base classifiers as a probability mass function of a binomial distribution:

$$P(y \geq k) = \sum_k^n \binom{n}{k} \epsilon^k (1 - \epsilon)^{n-k} = \epsilon_{\text{ensemble}}$$

Here, $\binom{n}{k}$ is the binomial coefficient n choose k . In other words, we compute the probability that the prediction of the ensemble is wrong. Now let's take a look

at a more concrete example of 11 base classifiers ($n = 11$) with an error rate of 0.25 ($\epsilon = 0.25$):

$$P(y \geq k) = \sum_{k=6}^{11} \binom{11}{k} 0.25^k (1 - \epsilon)^{11-k} = 0.034$$

7.2 Implementing a simple majority vote classifier

Our goal is to build a stronger meta-classifier that balances out the individual classifiers' weaknesses on a particular dataset. In more precise mathematical terms, we can write the weighted majority vote as follows:

$$\hat{y} = \arg \max_i \sum_{j=1}^m w_j \chi_A(C_j(\mathbf{x}) = i)$$

Let's assume that we have an ensemble of three base classifiers $C_j (j \in 0, 1)$ and want to predict the class label of a given sample instance \mathbf{x} . Two out of three base classifiers predict the class label 0, and one C_3 predicts that the sample belongs to class 1. If we weight the predictions of each base classifier equally, the majority vote will predict that the sample belongs to class 0:

$$C_1(\mathbf{x}) \rightarrow 0, C_2(\mathbf{x}) \rightarrow 0, C_3(\mathbf{x}) \rightarrow 1$$

$$\hat{y} = \text{mode}\{0, 0, 1\} = 0$$

Now let's assign a weight of 0.6 to C_3 and weight C_1 and C_2 by a coefficient of 0.2, respectively.

$$\begin{aligned} \hat{y} &= \arg \max_i \sum_{j=1}^m w_j \chi_A(C_j(\mathbf{x}) = i) \\ &= \arg \max_i [0.2 \times i_0 + 0.2 \times i_0 + 0.6 \times i_1] = 1 \end{aligned}$$

More intuitively, since $3 \times 0.2 = 0.6$, we can say that the prediction made by C_3 has three times more weight than the predictions by C_1 or C_2 , respectively. We can write this as follows:

$$\hat{y} = \text{mode}\{0, 0, 1, 1, 1\} = 1$$

[...] The modified version of the majority vote for predicting class labels from probabilities can be written as follows:

$$\hat{y} = \arg \max_i \sum_{j=1}^m w_j p_{ij}$$

Here, p_{ij} is the predicted probability of the j th classifier for class label i . To continue with our previous example, let's assume that we have a binary classification problem with class labels $i \in \{0, 1\}$ and an ensemble of three classifiers $C_j (j \in \{1, 2, 3\})$. Let's assume that the classifier C_j returns the following class membership probabilities for a particular sample \mathbf{x} :

$$C_1(\mathbf{x}) \rightarrow [0.9, 0.1], C_2(\mathbf{x}) \rightarrow [0.8, 0.2], C_3(\mathbf{x}) \rightarrow [0.4, 0.6]$$

We can then calculate the individual class probabilities as follows:

$$p(i_0|\mathbf{x}) = 0.2 \times 0.9 + 0.2 \times 0.8 + 0.6 \times 0.4 = 0.58$$

$$p(i_1|\mathbf{x}) = 0.2 \times 0.1 + 0.2 \times 0.2 + 0.6 \times 0.06 = 0.42$$

$$\hat{y} = \arg \max_i [p(i_0|\mathbf{x}), p(i_1|\mathbf{x})] = 0$$

7.2.1 Combining different algorithms for classification with majority vote

7.3 Evaluating and tuning the ensemble classifier

7.4 Bagging – building an ensemble of classifiers from bootstrap samples

7.5 Leveraging weak learners via adaptive boosting

[...] The original boosting procedure is summarized in four key steps as follows:

1. Draw a random subset of training samples d_1 without replacement from the training set D to train a weak learner C_1 .
2. Draw second random training subset d_2 without replacement from the training set and add 50 percent of the samples that were previously misclassified to train a weak learner C_2 .
3. Find the training samples d_3 in the training set D on which C_1 and C_2 disagree to train a third weak learner C_3
4. Combine the weak learners C_1, C_2 , and C_3 via majority voting.

[...] Now that have a better understanding behind the basic concept of AdaBoost, let's take a more detailed look at the algorithm using pseudo code. For clarity, we will denote element-wise multiplication by the cross symbol (\times) and the dot product between two vectors by a dot symbol (\cdot), respectively. The steps are as follows:

1. Set weight vector \mathbf{w} to uniform weights where $\sum_i w_i = 1$.
2. For j in m boosting rounds, do the following:
 - (a) Train a weighted weak learner: $C_j = \text{train}(\mathbf{X}, \mathbf{y}, \mathbf{w})$.
 - (b) Predict class labels: $\hat{\mathbf{y}} = \text{predict}(C_j, \mathbf{X})$.
 - (c) Compute the weighted error rate: $\epsilon = \mathbf{w} \cdot (\hat{\mathbf{y}} \neq \mathbf{y})$.
 - (d) Compute the coefficient α_j : $\alpha_j = 0.5 \log \frac{1-\epsilon}{\epsilon}$.
 - (e) Update the weights: $\mathbf{w} := \mathbf{w} \times \exp(-\alpha_j \times \hat{\mathbf{y}} \times \mathbf{y})$.
 - (f) Normalize weights to sum to 1: $\mathbf{w} := \mathbf{w} / \sum_i w_i$.
3. Compute the final prediction: $\hat{\mathbf{y}} = (\sum_{j=1}^m (\alpha_j \times \text{predict}(C_j, \mathbf{X})) > 0)$.

Note that the expression $(\hat{\mathbf{y}} \neq \mathbf{y})$ in step 5 refers to a vector of 1s and 0s, where a 1 is assigned if the prediction is incorrect and 0 is assigned otherwise.

Sample indices	x	y	Weights	$\hat{y}(x \leq 3.0)?$	Correct?	Updated weights
1	1.0	1	0.1	1	Yes	0.072
2	2.0	1	0.1	1	Yes	0.072
3	3.0	1	0.1	1	Yes	0.072
4	4.0	-1	0.1	-1	Yes	0.072
5	5.0	-1	0.1	-1	Yes	0.072
6	6.0	-1	0.1	-1	Yes	0.072
7	7.0	1	0.1	-1	No	0.167
8	8.0	1	0.1	-1	No	0.167
9	9.0	1	0.1	-1	No	0.167
10	10.0	-1	0.1	-1	Yes	0.072

Since the computation of the weight updates may look a little bit complicated at first, we will now follow the calculation step by step. We start by computing the weighted error rate ϵ as described in step 5:

$$\begin{aligned}
 \epsilon &= 0.1 \times 0 + 0.1 \times 0 + 0.1 \times 0 + 0.1 \times 0 + 0.1 \times 0 + 0.1 \times 0 + 0.1 \times 1 + 0.1 \times 1 + 0.1 \times 1 + 0.1 \times 0 \\
 &= \frac{3}{10} = 0.3
 \end{aligned}$$

Next we compute the coefficient α_j (shown in step 6), which is later used in step 7 to update the weights as well as for the weights in majority vote prediction (step 10):

$$\alpha_j = 0.5 \log \left(\frac{1 - \epsilon}{\epsilon} \right) \approx 0.424$$

After we have computed the coefficient α_j we can now update the weight vector using the following equation:

$$\mathbf{w} := \mathbf{w} \times \exp(-\alpha_j \times \hat{\mathbf{y}} \times \mathbf{y})$$

Here, $\hat{\mathbf{y}} \times \mathbf{y}$ is an element-wise multiplication between the vectors of the predicted and true class labels, respectively. Thus, if a prediction \hat{y}_i is correct, $\hat{y}_i \times y_i$ will have a positive sign so that we decrease the i th weight since α_j is a positive number as well:

$$0.1 \times \exp(-0.424 \times 1 \times 1) \approx 0.065$$

Similarly, we will increase the i th weight if \hat{y}_i predicted the label incorrectly like this:

$$0.1 \times \exp(-0.424 \times 1 \times (-1)) \approx 0.153$$

Or like this:

$$0.1 \times \exp(-0.424 \times (-1) \times 1) \approx 0.153$$

After we update each weight in the weight vector, we normalize the weights so that they sum up to 1 (step 8):

$$\mathbf{w} := \frac{\mathbf{w}}{\sum_i w_i}$$

Here, $\sum_i w_i = 7 \times 0.065 + 3 \times 0.153 = 0.914$.

Thus, each weight that corresponds to a correctly classified sample will be reduced from the initial value of 0.1 to $0.065/0.914 \approx 0.071$ for the next round of boosting. Similarly, the weights of each incorrectly classified sample will increase from 0.1 to $0.153/0.914 \approx 0.167$.

7.6 Summary

Chapter 8

Applying Machine Learning to Sentiment Analysis

8.1 Obtaining the IMDb movie review dataset

8.2 Introducing the bag-of-words model

8.2.1 Transforming words into feature vectors

8.2.2 Assessing word relevancy via term frequency-inverse document frequency

The *tf-idf* can be defined as the product of the *term frequency* and the *inverse document frequency*:

$$\text{tf-idf}(t, d) = \text{tf}(t, d) \times \text{idf}(t, d)$$

Here the $\text{tf}(t, d)$ is the term frequency that we introduced in the previous section, and the inverse document frequency $\text{idf}(t, d)$ can be calculated as:

$$\text{idf}(t, d) = \log \frac{n_d}{1 + \text{df}(d, t)},$$

where n_d is the total number of documents, and $\text{df}(d, t)$ is the number of documents d that contain the term t . Note that adding the constant 1 to the denominator is optional and serves the purpose of assigning a non-zero value to terms that occur in all training samples; the log is used to ensure that low document frequencies are not given too much weight.

However, if we'd manually calculated the *tf-idfs* of the individual terms in our feature vectors, we'd have noticed that the *TfidfTransformer* calculates the *tf-idfs* slightly differently compared to the standard textbook equations that we defined earlier. The equations for the *idf* and *tf-idf* that were implemented in *scikit-learn* are:

$$\text{idf}(t, d) = \log \frac{1 + n_d}{1 + \text{df}(d, t)}$$

The tf-idf equation that was implemented in scikit-learn is as follows:

$$\text{tf-idf}(t, d) = \text{tf}(t, d) \times (\text{idf}(t, d) + 1).$$

While it is also more typical to normalize the raw term frequencies before calculating the tf-idfs, the *TfidfTransformer* normalizes the tf-idfs directly. By default (norm='l2'), scikit-learn's *TfidfTransformer* applies the L2-normalization, which returns a vector of length 1 by dividing an un-normalized feature vector \mathbf{v} by its L2-norm:

$$\mathbf{v}_{\text{norm}} = \frac{\mathbf{v}}{\|\mathbf{v}\|_2} = \frac{\mathbf{v}}{\sqrt{v_1^2 + v_2^2 + \dots + v_n^2}} = \frac{\mathbf{v}}{(\sum_{i=1}^n v_i^2)^{1/2}}$$

8.2.3 Cleaning text data

8.2.4 Processing documents into tokens

8.3 Training a logistic regression model for document classification

8.4 Working with bigger data ? online algorithms and out-of-core learning

8.5 Summary

... to be continued ...