LIBSVM: a Library for Support Vector Machines

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Abstract

LIBSVM is a library for support vector machines (SVM). Its goal is to help users to easily use SVM as a tool. In this document, we present all its implementation details. For the use of LIBSVM, the README file included in the package and the LIBSVM FAQ provide the information.

1 Introduction

LIBSVM is a library for support vector classification (SVM) and regression. Its goal is to let users can easily use SVM as a tool. In this document, we present all its implementation details. For the use of LIBSVM, the README file included in the package provides the information.

In Section 2, we show formulations used in LIBSVM: C-support vector classification (C-SVC), ν-support vector classification (ν-SVC), distribution estimation (one-class SVM), ε-support vector regression (ε-SVR), and ν-support vector regression (ν-SVR). We discuss the implementation of solving quadratic problems in Section 3. Section 4 describes two implementation techniques: shrinking and caching. Then in Section 5, we discuss the implementation of multi-class classification. We now also support different penalty parameters for unbalanced data. Details are in Section 6. Model selection is very important for obtaining the best generalization. LIBSVM provides simple and useful tools which are discussed in 7.

2 Formulations

2.1 C-Support Vector Classification (Binary Case)

Given training vectors $x_i \in \mathbb{R}^n, i = 1, \ldots, l$, in two classes, and a vector $y \in \mathbb{R}^l$ such that $y_i \in \{1, -1\}$, C-SVC (Cortes and Vapnik, 1995; Vapnik, 1998) solves the following primal problem:

$$\min_{w, b, \xi} \quad \frac{1}{2} w^T w + C \sum_{i=1}^{l} \xi_i$$

subject to \quad $y_i(w^T \phi(x_i) + b) \geq 1 - \xi_i,$ \quad $\xi_i \geq 0, i = 1, \ldots, l.$
Its dual is

\[
\min_{\alpha} \quad \frac{1}{2} \alpha^T Q \alpha - e^T \alpha \\
\text{subject to} \quad 0 \leq \alpha_i \leq C, \quad i = 1, \ldots, l,
\]

where \( e \) is the vector of all ones, \( C > 0 \) is the upper bound, \( Q \) is an \( l \) by \( l \) positive semidefinite matrix, \( Q_{ij} \equiv y_i y_j K(x_i, x_j) \), and \( K(x_i, x_j) \equiv \phi(x_i)^T \phi(x_j) \) is the kernel. Here training vectors \( x_i \) are mapped into a higher (maybe infinite) dimensional space by the function \( \phi \).

The decision function is

\[ \text{sgn}(\sum_{i=1}^{l} y_i \alpha_i K(x_i, x) + b). \]

### 2.2 \( \nu \)-Support Vector Classification (Binary Case)

The \( \nu \)-support vector classification (Schölkopf et al., 2000) uses a new parameter \( \nu \) which let one control the number of support vectors and errors. The parameter \( \nu \in (0,1] \) is an upper bound on the fraction of training errors and a lower bound of the fraction of support vectors.

Details of the algorithm implemented in LIBSVM can be found in (Chang and Lin, 2001). Given training vectors \( x_i \in \mathbb{R}^n \), \( i = 1, \ldots, l \), in two classes, and a vector \( y \in \mathbb{R}^l \) such that \( y_i \in \{1, -1\} \), the primal form considered is:

\[
\min_{w, b, \xi, \rho} \quad \frac{1}{2} w^T w - \nu \rho + \frac{1}{l} \sum_{i=1}^{l} \xi_i \\
\text{subject to} \quad y_i (w^T \phi(x_i) + b) \geq \rho - \xi_i, \\
\xi_i \geq 0, \quad i = 1, \ldots, l, \quad \rho \geq 0.
\]

The dual is:

\[
\min_{\alpha} \quad \frac{1}{2} \alpha^T Q \alpha \\
\text{subject to} \quad 0 \leq \alpha_i \leq 1/l, \quad i = 1, \ldots, l,
\]

where \( Q_{ij} \equiv y_i y_j K(x_i, x_j) \).

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The decision function is:

\[ \text{sgn}\left( \sum_{i=1}^{l} y_i \alpha_i (K(x_i, x) + b) \right). \]

In [Crisp and Burges 2000; Chang and Lin 2001], it has been shown that \( \mathbf{e}^T \alpha \geq \nu \) can be replaced by \( \mathbf{e}^T \alpha = \nu \). With this property, in LIBSVM, we solve a scaled version of (2.3):

\[
\begin{align*}
\min_{\alpha} & \quad \frac{1}{2} \alpha^T Q \alpha \\
\text{subject to} & \quad 0 \leq \alpha_i \leq 1, \quad i = 1, \ldots, l, \\
& \quad \mathbf{e}^T \alpha = \nu l, \\
& \quad y^T \alpha = 0.
\end{align*}
\]

We output \( \alpha/\rho \) so the computed decision function is:

\[ \text{sgn}\left( \sum_{i=1}^{l} y_i (\alpha_i/\rho) (K(x_i, x) + b) \right) \]

and then two margins are

\[ y_i (\mathbf{w}^T \phi(x_i) + b) = \pm 1 \]

which are the same as those of C-SVC.

### 2.3 Distribution Estimation (One-class SVM)

One-class SVM was proposed by [Schölkopf et al. 2001] for estimating the support of a high-dimensional distribution. Given training vectors \( x_i \in \mathbb{R}^n, i = 1, \ldots, l \) without any class information, the primal form in [Schölkopf et al. 2001] is:

\[
\begin{align*}
\min_{\mathbf{w}, b, \xi, \rho} & \quad \frac{1}{2} \mathbf{w}^T \mathbf{w} - \rho + \frac{1}{\nu l} \sum_{i=1}^{l} \xi_i \\
\text{subject to} & \quad \mathbf{w}^T \phi(x_i) \geq \rho - \xi_i, \\
& \quad \xi_i \geq 0, i = 1, \ldots, l.
\end{align*}
\]

The dual is:

\[
\begin{align*}
\min_{\alpha} & \quad \frac{1}{2} \alpha^T Q \alpha \\
\text{subject to} & \quad 0 \leq \alpha_i \leq 1/(\nu l), i = 1, \ldots, l, \quad (2.4) \\
& \quad \mathbf{e}^T \alpha = 1,
\end{align*}
\]

where \( Q_{ij} = K(x_i, x_j) \equiv \phi(x_i)^T \phi(x_j) \).
In LIBSVM we solve a scaled version of (2.4):

$$\min \frac{1}{2} \alpha^T Q \alpha$$
subject to $0 \leq \alpha_i \leq 1$, $i = 1, \ldots, l$,
$$e^T \alpha = \nu l.$$  

The decision function is
$$\text{sgn}(\sum_{i=1}^{l} \alpha_i K(x_i, x) - \rho).$$

### 2.4 $\epsilon$-Support Vector Regression ($\epsilon$-SVR)

Given a set of data points, $\{(x_1, z_1), \ldots, (x_l, z_l)\}$, such that $x_i \in R^n$ is an input and $z_i \in R^1$ is a target output, the standard form of support vector regression (Vapnik, 1998) is:

$$\min_{w, b, \xi, \xi^*} \frac{1}{2} w^T w + C \sum_{i=1}^{l} \xi_i + C \sum_{i=1}^{l} \xi_i^*$$
subject to
$$w^T \phi(x_i) + b - z_i \leq \epsilon + \xi_i,$$
$$z_i - w^T \phi(x_i) - b \leq \epsilon + \xi_i^*,$$
$$\xi_i, \xi_i^* \geq 0, i = 1, \ldots, l.$$

The dual is:

$$\min_{\alpha, \alpha^*} \frac{1}{2} (\alpha - \alpha^*)^T Q (\alpha - \alpha^*) + \epsilon \sum_{i=1}^{l} (\alpha_i + \alpha_i^*) + \sum_{i=1}^{l} z_i (\alpha_i - \alpha_i^*)$$
subject to
$$\sum_{i=1}^{l} (\alpha_i - \alpha_i^*) = 0, 0 \leq \alpha_i, \alpha_i^* \leq C, i = 1, \ldots, l.$$  

where $Q_{ij} = K(x_i, x_j) \equiv \phi(x_i)^T \phi(x_j)$.

The approximate function is:
$$\sum_{i=1}^{l} (-\alpha_i + \alpha_i^*) K(x_i, x) + b.$$

### 2.5 $\nu$-Support Vector Regression ($\nu$-SVR)

Similar to $\nu$-SVC, for regression, (Schölkopf et al., 2000) use a parameter $\nu$ to control the number of support vectors. However, unlike $\nu$-SVC where $C$ is replaced by $\nu$ here $\nu$ replaces...
the parameter $\epsilon$ of $\epsilon$-SVR. The primal form is

$$\min_{w,b,\xi,\epsilon} \frac{1}{2}w^T w + C(\nu \epsilon + \frac{1}{l}\sum_{i=1}^{l}(\xi_i + \xi_i^*))$$  \hspace{1cm} (2.6)$$
subject to

$$(w^T \phi(x_i) + b) - z_i \leq \epsilon + \xi_i,$$

$$z_i - (w^T \phi(x_i) + b) \leq \epsilon + \xi_i^*,$$

$$\xi_i, \xi_i^* \geq 0, i = 1, \ldots, l, \epsilon \geq 0.$$  

and the dual is

$$\min_{\alpha, \alpha^*} \frac{1}{2}(\alpha - \alpha^*)^T Q(\alpha - \alpha^*) + z^T (\alpha - \alpha^*)$$
subject to

$$e^T (\alpha - \alpha^*) = 0, \ e^T (\alpha + \alpha^*) \leq C\nu,$$

$$0 \leq \alpha_t, \alpha_t^* \leq C/l, \ i = 1, \ldots, l.$$  \hspace{1cm} (2.7)$$

Similarly, the inequality $e^T (\alpha + \alpha^*) \leq C\nu$ can be replaced by an equality. In LIBSVM, we consider $C \leftarrow C/l$ so the dual problem solved is:

$$\min_{\alpha, \alpha^*} \frac{1}{2}(\alpha - \alpha^*)^T Q(\alpha - \alpha^*) + z^T (\alpha - \alpha^*)$$
subject to

$$e^T (\alpha - \alpha^*) = 0, \ e^T (\alpha + \alpha^*) = C\nu,$$

$$0 \leq \alpha_t, \alpha_t^* \leq C, \ i = 1, \ldots, l.$$  \hspace{1cm} (2.8)$$

Then the decision function is

$$\sum_{i=1}^{l}(-\alpha_i + \alpha_i^*)K(x_i, x) + b,$$
the same as that of $\epsilon$-SVR.

3 Solving the Quadratic Problems

3.1 The Decomposition Method for $C$-SVC, $\epsilon$-SVR, and One-class SVM

We consider the following general form of $C$-SVC, $\epsilon$-SVR, and one-class SVM:

$$\min_{\alpha} \frac{1}{2}\alpha^T Q\alpha + p^T \alpha$$
subject to

$$y^T \alpha = \Delta,$$

$$0 \leq \alpha_t \leq C, t = 1, \ldots, l.$$  \hspace{1cm} (3.1)$$
where $y_t = \pm 1, t = 1, \ldots, l$. It can be clearly seen that C-SVC and one-class SVM are already in the form of (3.1). For $\epsilon$-SVR, we consider the following reformulation of (2.5):

$$
\min_{\alpha, \alpha^*} \frac{1}{2}[\alpha^T, (\alpha^*)^T] \begin{bmatrix} Q & -Q \\ -Q & Q \end{bmatrix} \begin{bmatrix} \alpha \\ \alpha^* \end{bmatrix} + \begin{bmatrix} \epsilon e^T + z^T, \epsilon e^T - z^T \end{bmatrix}_T \begin{bmatrix} \alpha \\ \alpha^* \end{bmatrix}
$$

subject to

$$
y^T \begin{bmatrix} \alpha \\ \alpha^* \end{bmatrix} = 0, 0 \leq \alpha_t, \alpha^*_t \leq C, t = 1, \ldots, l,
$$

(3.2)

where $y$ is a $2l$ by 1 vector with $y_t = 1, t = 1, \ldots, l$ and $y_t = -1, t = l + 1, \ldots, 2l$.

The difficulty of solving (3.1) is the density of $Q$ because $Q_{ij}$ is in general not zero. In LIBSVM, we consider the decomposition method to conquer this difficulty. Some work on this method are, for example, (Osuna et al., 1997b; Joachims, 1998; Platt, 1998; Saunders et al., 1998).

**Algorithm 1 (Decomposition method)**

1. Given a number $q \leq l$ as the size of the working set. Find $\alpha^1$ as the initial solution. Set $k = 1$.

2. If $\alpha^k$ is an optimal solution of (2.2), stop. Otherwise, find a working set $B \subset \{1, \ldots, l\}$ whose size is $q$. Define $N \equiv \{1, \ldots, l\} \setminus B$ and $\alpha^k_B$ and $\alpha^k_N$ to be sub-vectors of $\alpha^k$ corresponding to $B$ and $N$, respectively.

3. Solve the following sub-problem with the variable $\alpha_B$:

$$
\min_{\alpha_B} \frac{1}{2} \alpha^T B Q B \alpha_B + (p_B + Q_B N \alpha^k_N)^T \alpha_B
$$

subject to

$$
y_B^T \alpha_B = \Delta - y_N^T \alpha^k_N,
$$

(3.3)

where $\begin{bmatrix} Q_{BB} & Q_{BN} \\ Q_{NB} & Q_{NN} \end{bmatrix}$ is a permutation of the matrix $Q$.

4. Set $\alpha_B^{k+1}$ to be the optimal solution of (3.3) and $\alpha_N^{k+1} \equiv \alpha_N^k$. Set $k \leftarrow k + 1$ and goto Step 2.

The basic idea of the decomposition method is that in each iteration, the indices $\{1, \ldots, l\}$ of the training set are separated to two sets $B$ and $N$, where $B$ is the working set and $N = \{1, \ldots, l\} \setminus B$. The vector $\alpha_N$ is fixed so the objective value becomes $\frac{1}{2} \alpha_B^T Q_{BB} \alpha_B - (p_B - Q_{BN} \alpha_N)^T \alpha_B + \frac{1}{2} \alpha_N^T Q_{NN} \alpha_N - p_N^T \alpha_N$. Then a sub-problem with the variable $\alpha_B$, i.e. (3.3), is solved. Note that $B$ is updated in each iteration. To simplify the notation, we simply use $B$ instead of $B^k$. The strict decrease of the objective function holds and the theoretical convergence will be discussed in Section 3.4.
3.2 Working Set Selection and Stopping Criteria for C-SVC, \(\epsilon\)-SVR, and One-class SVM

An important issue of the decomposition method is the selection of the working set \(B\). The Karush-Kuhn-Tucker (KKT) condition of (3.1) shows that there is a scalar \(b\) and two nonnegative vectors \(\lambda\) and \(\mu\) such that

\[
Q\alpha + p + by = \lambda - \mu,
\]

\[
\lambda_i\alpha_i = 0, \mu_i(C - \alpha)_i = 0,
\]

\[
\lambda_i \geq 0, \mu_i \geq 0, i = 1, \ldots, l.
\]

Note that if we write down the KKT conditions of both primal and dual forms, they are the same and the Lagrange multiplier of the linear constraint \(y^T\alpha = \Delta\) coincides with the bias term of the decision function. Therefore, the \(b\) here is the same as that in the decision function.

Then (3.4) can be rewritten as

\[
(Q\alpha + p + by)_i \geq 0 \text{ if } \alpha_i = 0,
\]

\[
= 0 \text{ if } 0 < \alpha_i < C,
\]

\[
\leq 0 \text{ if } \alpha_i = C.
\]

By using \(y_i = \pm 1, i = 1, \ldots, l\) and the assumption that \(C > 0\), this further implies that

\[
y_t = 1, \alpha_t < C \quad \Rightarrow \quad (Q\alpha + p)_t + b \geq 0 \quad \Rightarrow \quad b \geq -(Q\alpha + p)_t = -\nabla f(\alpha)_t,
\]

\[
y_t = -1, \alpha_t > 0 \quad \Rightarrow \quad (Q\alpha + p)_t - b \leq 0 \quad \Rightarrow \quad b \geq (Q\alpha + p)_t = \nabla f(\alpha)_t,
\]

\[
y_t = -1, \alpha_t < C \quad \Rightarrow \quad (Q\alpha + p)_t - b \leq 0 \quad \Rightarrow \quad b \leq -(Q\alpha + p)_t = -\nabla f(\alpha)_t,
\]

\[
y_t = 1, \alpha_t > 0 \quad \Rightarrow \quad (Q\alpha + p)_t + b \leq 0 \quad \Rightarrow \quad b \leq -(Q\alpha + p)_t = -\nabla f(\alpha)_t,
\]

where \(f(\alpha) \equiv \frac{1}{2}\alpha^TQ\alpha + p^T\alpha\) and \(\nabla f(\alpha)\) is the gradient of \(f(\alpha)\) at \(\alpha\). We consider

\[
i \equiv \arg \max(\{-\nabla f(\alpha)_t \mid y_t = 1, \alpha_t < C\}, \{\nabla f(\alpha)_t \mid y_t = -1, \alpha_t > 0\}),
\]

\[
j \equiv \arg \min(\{\nabla f(\alpha)_t \mid y_t = -1, \alpha_t < C\}, \{-\nabla f(\alpha)_t \mid y_t = 1, \alpha_t > 0\}).
\]

We then use \(B = \{i, j\}\) as the working set for the sub-problem (3.3) of the decomposition method. Here \(i\) and \(j\) are the two elements which violate the KKT condition the most.

The idea of using only two elements for the working set are from the Sequential Minimal Optimization (SMO) by (Platt, 1998). The main advantage is that an analytic solution of (3.3) can be obtained so there is no need to use an optimization software. Note that (3.6) and (3.7) are a special case of the working set selection of the software \(SVMlight\) by (Joachims, 1998). To be more precise, in \(SVMlight\), if \(\alpha\) is the current solution, the
The following problem is solved:

$$\min_d \nabla f(\alpha)^T d$$

subject to

$$y^T d = 0, -1 \leq d_t \leq 1,$$
$$d_t \geq 0, \text{ if } \alpha_t = 0, d_t \leq 0, \text{ if } \alpha_t = C,$$
$$|\{d_t \mid d_t \neq 0\}| = q.$$  \hspace{1cm} (3.8)

Note that \(|\{d_t \mid d_t \neq 0\}|\) means the number of components of \(d\) which are not zero. The constraint (3.9) implies that a descent direction involving only \(q\) variables is obtained. Then components of \(\alpha\) with non-zero \(d_t\) are included in the working set \(B\) which is used to construct the sub-problem (3.3). Note that \(d\) is only used for identifying \(B\) but not as a search direction.

It can be clearly seen that if \(q = 2\), the solution of (3.8) is

$$i = \arg \min \{\nabla f(\alpha)_i d_t \mid y_i d_t = 1; d_t \geq 0, \text{ if } \alpha_t = 0; d_t \leq 0, \text{ if } \alpha_t = C.\},$$
$$j = \arg \min \{\nabla f(\alpha)_j d_t \mid y_j d_t = -1; d_t \geq 0, \text{ if } \alpha_t = 0; d_t \leq 0, \text{ if } \alpha_t = C.\},$$

which is the same as (3.6) and (3.7). We also notice that this is also the modification 2 of the algorithms in (Keerthi et al., 2001).

We then define

$$g_i = \begin{cases} -\nabla f(\alpha)_i & \text{if } y_i = 1, \alpha_i < C, \\ \nabla f(\alpha)_i & \text{if } y_i = -1, \alpha_i > 0, \end{cases}$$  \hspace{1cm} (3.10)

and

$$g_j = \begin{cases} -\nabla f(\alpha)_j & \text{if } y_j = -1, \alpha_j < C, \\ \nabla f(\alpha)_j & \text{if } y_j = 1, \alpha_j > 0. \end{cases}$$  \hspace{1cm} (3.11)

From (3.5), we know

$$g_i \leq -g_j$$  \hspace{1cm} (3.12)

implies that \(\alpha\) is an optimal solution of (2.2). Practically the stopping criteria can be written and implemented as:

$$g_i \leq -g_j + \epsilon,$$  \hspace{1cm} (3.13)

where \(\epsilon\) is a small positive number.

### 3.3 A Note on the Working Set Selection of \(\epsilon\)-SVR

The \(\epsilon\)-SVR formulation involves with two closely related sets of variables: \(\alpha\) and \(\alpha^*\). An important property is that for any optimal solution, \(\alpha_i \alpha_i^* = 0, i = 1, \ldots, l\). In addition, this is true is for the zero initial solution. Therefore, to maintain this property throughout
all iterations so the number of nonzero variables during iterations can be kept small, many existing decomposition methods for regression use pairs of indices from these two sets as the working set. For example, if \( \{\alpha_i, \alpha^*_j\} \) are chosen first from the violation of KKT discussed in Section 3.2, they include \( \{\alpha^*_i, \alpha_j\} \) into the working set. Then a sub-problem of four variables \((\alpha_i, \alpha^*_i, \alpha_j, \alpha^*_j)\) is solved.

Using pairs of variables as the working set makes the implementation different from that for support vector classification. We tend to avoid this as we hope to have a simple software. In addition, a larger optimization sub-problem has to be solved in each iteration. However, from Lin (2001b, Theorem 4.1), it has been shown that if we use only the two variables selected by the method in Section 3.2, the property \( \alpha_i \alpha^*_i = 0, i = 1, \ldots, l \) still holds throughout all iterations. Then in Liao et al. (2002) we show that even if we expand the working set and sub-problem to have four variables, in most decomposition iterations, only the two originally selected variables are changed. In other words, no matter we expand the working set to be pairs of variables or not, the number of iterations is about the same. Therefore, in our implementation we do not consider \( \epsilon \)-SVR as a different from from C-SVR so the same decomposition method is applied.

3.4 Convergence of the Decomposition Method

The convergence of decomposition methods was first studied in Chang et al. (2000) but algorithms discussed there do not coincide with existing implementations. In this section we will discuss only convergence results related to the specific decomposition method in Section 3.2.

From Keerthi and Gilbert (2002) we have

**Theorem 3.1** Given any \( \epsilon > 0 \), after a finite number of iterations (3.13) will be satisfied.

This theorem establishes the so-called “finite termination” property so we are sure that after finite steps the algorithm will stop.

For asymptotic convergence, from Lin (2002a), we have

**Theorem 3.2** If \( \{\alpha^k\} \) is the sequence generated by the decomposition method in Section 3.2, the limit of any its convergent subsequence is an optimal solution of (3.1).

Note that Theorem 3.1 does not imply Theorem 3.2 as if we consider \( g_i \) and \( g_j \) in (3.13) as functions of \( \alpha \), they are not continuous. Hence we cannot take limit on both sides of (3.13) and claim that any convergent point has already satisfies the KKT condition.
Theorem 3.2 was first proved as a special case of general results in [Lin, 2001b] where some assumptions are needed. Now the proof in [Lin, 2002a] does not require any assumption.

For local convergence, as the algorithm used here is a special case of the one discussed in [Lin, 2001a], we have the following theorem

**Theorem 3.3** If $Q$ is positive definite and the dual optimization problem is degenerate (see Assumption 2 in [Lin, 2001a]), then there is $c < 1$ such that after $k$ is large enough,

$$f(\alpha^{k+1}) - f(\alpha^*) \leq c(f(\alpha^k) - f(\alpha^*)),$$

where $\alpha^*$ is the optimal solution of (3.1).

That is, LIBSVM is linearly convergent.

All the above results are for “valid kernels” which can be considered as the inner product of two feature vectors. Therefore, $Q$ is positive semi-definite. For some kernels such as the sigmoid kernel $(K(x_i, x_j) = \tanh(\alpha x_i^T x_j - r))$, $Q$ may not be positive semi-definite. Therefore, the quadratic problem (3.1) is not convex so may have several local minima. Using a slightly modified Algorithm I, LIBSVM still guarantees to converges to local minima. More details are in [Lin and Lin, 2003].

In addition, the discussion here is about asymptotic convergence. We investigate the computational complexity in Section 4.3.

### 3.5 The Decomposition Method for $\nu$-SVC and $\nu$-SVR

Both $\nu$-SVC and $\nu$-SVR can be considered as the following general form:

$$\begin{align*}
\min_{\alpha} & \quad \frac{1}{2} \alpha^T Q \alpha + p^T \alpha \\
\text{subject to} & \quad y^T \alpha = \Delta_1, \\
& \quad e^T \alpha = \Delta_2,
\end{align*}$$

(3.14)

0 \leq \alpha_t \leq C, t = 1, \ldots, l.

The decomposition method is the same as Algorithm I but the sub-problem is different:

$$\begin{align*}
\min_{\alpha_B} & \quad \frac{1}{2} \alpha_B^T Q_B \alpha_B + (p_B + Q_{BN} \alpha_N^k)^T \alpha_B \\
\text{subject to} & \quad y_B^T \alpha_B = \Delta_1 - y_N^T \alpha_N^k, \\
& \quad e_B^T \alpha_B = \Delta_2 - e_N^T \alpha_N^k,
\end{align*}$$

(3.15)

0 \leq (\alpha_B)_t \leq C, t = 1, \ldots, q.
Now if only two elements $i$ and $j$ are selected but $y_i \neq y_j$, then $y_B^T \alpha_B = \Delta_1 - y_N^T \alpha_N^k$ and $e_B^T \alpha_B = \Delta_2 - e_N^T \alpha_N^k$ imply that there are two equations with two variables so (3.15) has only one feasible point. Therefore, from $\alpha^k$, the solution cannot be moved any more.

On the other hand, if $y_i = y_j$, $y_B^T \alpha_B = \Delta_1 - y_N^T \alpha_N^k$ and $e_B^T \alpha_B = \Delta_2 - e_N^T \alpha_N^k$ become the same equality so there are multiple feasible solutions. Therefore, we have to keep $y_i = y_j$ while selecting the working set.

The KKT condition of (3.14) shows

$$\nabla f(\alpha)_i - \rho + by_i = \begin{cases} 0 & \text{if } 0 < \alpha_i < C, \\ \geq 0 & \text{if } \alpha_i = 0, \\ \leq 0 & \text{if } \alpha_i = C. \end{cases}$$

Define

$$r_1 \equiv \rho - b, \ r_2 \equiv \rho + b.$$ 

If $y_i = 1$ the KKT condition becomes

$$\nabla f(\alpha)_i - r_1 \begin{cases} \geq 0 & \text{if } \alpha_i < C, \\ \leq 0 & \text{if } \alpha_i > 0. \end{cases} \tag{3.16}$$

On the other hand, if $y_i = -1$, it is

$$\nabla f(\alpha)_i - r_2 \begin{cases} \geq 0 & \text{if } \alpha_i < C, \\ \leq 0 & \text{if } \alpha_i > 0. \end{cases} \tag{3.17}$$

Hence, indices $i$ and $j$ are selected from either

$$i = \arg\min_t \{ \nabla f(\alpha)_t | y_t = 1, \alpha_t < C \},$$
$$j = \arg\max_t \{ \nabla f(\alpha)_t | y_t = 1, \alpha_t > 0 \}, \tag{3.18}$$

or

$$i = \arg\min_t \{ \nabla f(\alpha)_t | y_t = -1, \alpha_t < C \},$$
$$j = \arg\max_t \{ \nabla f(\alpha)_t | y_t = -1, \alpha_t > 0 \}, \tag{3.19}$$

depending on which one gives a smaller $\nabla f(\alpha)_i - \nabla f(\alpha)_j$ (i.e. larger KKT violations).

This was first proposed in [Keerthi and Gilbert, 2002]. Some details for classification and regression can be seen in [Chang and Lin, 2001, Section 4] and [Chang and Lin, 2002], respectively.

The stopping criterion will be described in Section 3.7.
3.6 Analytical Solutions

Now (3.3) is a simple problem with only two variables:

$$\min_{\alpha_i, \alpha_j} \frac{1}{2} \begin{bmatrix} \alpha_i & \alpha_j \end{bmatrix} \begin{bmatrix} Q_{ii} & Q_{ij} \\ Q_{ji} & Q_{jj} \end{bmatrix} \begin{bmatrix} \alpha_i \\ \alpha_j \end{bmatrix} + (Q_{i,N} \alpha_N - 1) \alpha_i + (Q_{j,N} \alpha_N - 1) \alpha_j$$

subject to

$$y_i\alpha_i + y_j\alpha_j = \Delta' \equiv \Delta - y_N^T \alpha_N^k,$$  (3.20)

$$0 \leq \alpha_i, \alpha_j \leq C.$$  

(Platt [1998]) substituted $$\alpha_i = y_i(\Delta - y_N^T \alpha_N - y_j\alpha_j)$$ into the objective function of (3.3) and solved an unconstrained minimization on $$\alpha_j$$. The following solution is obtained:

$$\alpha_j^{\text{new}} = \begin{cases} \alpha_j + \frac{-G_i - G_j}{Q_{ii} + Q_{jj} + 2Q_{ij}} & \text{if } y_i \neq y_j, \\ \alpha_j + \frac{G_i - G_j}{Q_{ii} + Q_{jj} - 2Q_{ij}} & \text{if } y_i = y_j, \end{cases}$$  (3.21)

where

$$G_i \equiv \nabla f(\alpha)_i \text{ and } G_j \equiv \nabla f(\alpha)_j.$$  

If this value is outside the possible region of $$\alpha_j$$ (that is, exceeds the feasible region of (3.3)), the value of (3.21) is clipped into the feasible region and is assigned as the new $$\alpha_j$$. For example, if $$y_i = y_j$$ and $$C \leq \alpha_i + \alpha_j \leq 2C$$, $$\alpha_j^{\text{new}}$$ must satisfy

$$L \equiv \alpha_i + \alpha_j - C \leq \alpha_j^{\text{new}} \leq C \equiv H$$

as the largest value $$\alpha_i^{\text{new}}$$ and $$\alpha_j^{\text{new}}$$ can be is $$C$$. Hence if

$$\alpha_j + \frac{G_i - G_j}{Q_{ii} + Q_{jj} + 2Q_{ij}} \leq L,$$

we define $$\alpha_j^{\text{new}} \equiv L$$ and then

$$\alpha_i^{\text{new}} = \alpha_i + \alpha_j - \alpha_j^{\text{new}} = C.$$  (3.22)

This can be illustrated by Figure 1 where it is like we are optimizing a quadratic function over a line segment. The line segment is the intersection between the linear constraint $$y_i\alpha_i + y_j\alpha_j = \Delta'$$ and bounded constraints $$0 \leq \alpha_i, \alpha_j \leq C$$.

However, numerically the last equality of (3.22) may not hold. The floating-point operation will cause that

$$\alpha_i + \alpha_j - \alpha_j^{\text{new}}$$

$$= \alpha_i + \alpha_j - (\alpha_i + \alpha_j - C)$$

$$\neq C.$$
Therefore, in most SVM software, a small tolerance $\epsilon_a$ is specified and all $\alpha_i \geq C - \epsilon_a$ are considered to be at the upper bound and all $\alpha_i \leq \epsilon_a$ are considered to be zero. This is necessary as otherwise some data will be wrongly considered as support vectors. In addition, the calculation of the bias term $b$ also need correct identification of those $\alpha_i$ which are free (i.e. $0 < \alpha_i < C$).

In [Hsu and Lin 2002b], it has been pointed out that if all bounded $\alpha_i$ obtain their values using direct assignments, there is no need of using an $\epsilon_a$. To be more precise, for floating-point computation, if $\alpha_i \leftarrow C$ is assigned somewhere, a future floating-point comparison between $C$ and $C$ returns true as they both have the same internal representation. More details about its implementation will be presented in Section 6. Note that though this involves a little more operations, as solving the analytic solution of (3.20) takes only a small portion of the total computational time, the difference is negligible.

Another minor problem is that the denominator in (3.21) is sometime zero. When this happens,

$$Q_{ij} = \mp(Q_{ii} + Q_{jj})/2$$

so

$$Q_{ii}Q_{jj} - Q_{ij}^2$$

$$= Q_{ii}Q_{jj} - (Q_{ii} + Q_{jj})^2/4$$

$$= -(Q_{ii} - Q_{ij})^2/4 \leq 0.$$ 

Therefore, we know if $Q_{BB}$ is positive definite, the zero denominator in (3.21) never happens. Hence this problem happens only if $Q_{BB}$ is a 2 by 2 singular matrix. We discuss some situations where $Q_{BB}$ may be singular.

1. The function $\phi$ does not map data to independent vectors in a higher-dimensional space so $Q$ is only positive semidefinite. For example, using the linear or low-degree polynomial kernels. Then it is possible that a singular $Q_{BB}$ is picked.
2. Some kernels have a nice property that $\phi(x_i), i = 1, \ldots, l$ are independent if $x_i \neq x_j$. Thus $Q$ as well as all possible $Q_{BB}$ are positive definite. An example is the RBF kernel (Micchelli, 1986). However, for many practical data we have encountered, some of $x_i, i = 1, \ldots, l$ are the same. Therefore, several rows (columns) of $Q$ are exactly the same so $Q_{BB}$ may be singular.

However, even if the denominator of (3.21) is zero, there are no numerical problems: From (3.13), we note that

$$g_i + g_j \geq \epsilon$$

during the iterative process. Since

$$g_i + g_j = \pm(-G_i - G_j) \text{ if } y_i \neq y_j, \text{ and}$$
$$g_i + g_j = \pm(G_i - G_j) \text{ if } y_i = y_j,$$

the situation of $0/0$ which is defined as NaN by IEEE standard does not appear. Therefore, (3.21) returns $\pm\infty$ if the denominator is zero which can be detected as special quantity of IEEE standard and clipped to regular floating point number. However, due to the $\pm0$ defined under IEEE standard, we have to make sure that the denominator is $+0$ so the value goes to the desired $+\infty$ or $-\infty$.

If the kernel matrix is not positive semi-definite, $Q_{ii} + Q_{jj} + 2Q_{ij}$ (or $Q_{ii} + Q_{jj} - 2Q_{ij}$) may not be positive so (3.21) may not produce an update so the objective value is decreased. Furthermore, the algorithm may stay at the same point so end with an infinite loop. In (Lin and Lin, 2003), we have studied this issue in detail and proposed the following modification:

$$Q^\text{new} = \begin{cases} 
\alpha_j + \frac{-G_i - G_j}{\max(Q_{ii} + Q_{jj} + 2Q_{ij}, +0)} & \text{if } y_i \neq y_j, \\
\alpha_j + \frac{G_i - G_j}{\max(Q_{ii} + Q_{jj} - 2Q_{ij}, +0)} & \text{if } y_i = y_j.
\end{cases} \tag{3.23}
$$

This new formula guarantees the strict decrease of the objective function. Similar to the explanation when $Q_{ii} + Q_{jj} \pm 2Q_{ij} = 0$, we have to be careful about $\pm0$ in practical implementation. The implementation of (3.23) requires that max function to be defined so that $\max(-0, +0) = +0$ under the IEEE standard.

### 3.7 The Calculation of $b$ or $\rho$

After the solution $\alpha$ of the dual optimization problem is obtained, the variables $b$ or $\rho$ must be calculated as they are used in the decision function. Here we simply describe the case of $\nu$-SVC and $\nu$-SVR where $b$ and $\rho$ both appear. Other formulations are simplified cases of them.
The KKT condition of (3.14) has been shown in (3.16) and (3.17). Now we consider the case of $y_i = 1$. If there are $\alpha_i$ which satisfy $0 < \alpha_i < C$, then $r_1 = \nabla f(\alpha)_i$. Practically to avoid numerical errors, we average them:

$$r_1 = \frac{\sum_{0 < \alpha_i < C, y_i = 1} \nabla f(\alpha)_i}{\sum_{0 < \alpha_i < C, y_i = 1} 1}.$$ 

On the other hand, if there is no such $\alpha_i$, as $r_1$ must satisfy

$$\max_{\alpha_i = C, y_i = 1} \nabla f(\alpha)_i \leq r_1 \leq \min_{\alpha_i = 0, y_i = 1} \nabla f(\alpha)_i,$$

we take $r_1$ the midpoint of the range.

For $y_i = -1$, we can calculate $r_2$ in a similar way.

After $r_1$ and $r_2$ are obtained,

$$\rho = \frac{r_1 + r_2}{2} \quad \text{and} \quad b = \frac{r_1 - r_2}{2}.$$

Note that the KKT condition can be written as

$$\max_{\alpha_i > 0, y_i = 1} \nabla f(\alpha)_i \leq \min_{\alpha_i < C, y_i = 1} \nabla f(\alpha)_i$$

and

$$\max_{\alpha_i > 0, y_i = -1} \nabla f(\alpha)_i \leq \min_{\alpha_i < C, y_i = -1} \nabla f(\alpha)_i.$$ 

Hence practically we can use the following stopping criterion: The decomposition method stops if the iterate $\alpha$ satisfies the following condition:

$$\max(\ - \min_{\alpha_i < C, y_i = 1} \nabla f(\alpha)_i + \max_{\alpha_i > 0, y_i = 1} \nabla f(\alpha)_i, \quad \text{(3.24)}}$$

$$- \min_{\alpha_i < C, y_i = -1} \nabla f(\alpha)_i + \max_{\alpha_i > 0, y_i = -1} \nabla f(\alpha)_i < \epsilon, \quad \text{(3.25)}}$$

where $\epsilon > 0$ is a chosen stopping tolerance.

4 Shrinking and Caching

4.1 Shrinking

Since for many problems the number of free support vectors (i.e. $0 < \alpha_i < C$) is small, the shrinking technique reduces the size of the working problem without considering some bounded variables (Joachims [1998]). Near the end of the iterative process, the decomposition method identifies a possible set $A$ where all final free $\alpha_i$ may reside in. Indeed we can have the following theorem which shows that at the final iterations of the decomposition proposed in Section 3.2 only variables corresponding to a small set are still allowed to move (Lin [2002b], Theorem II.3):
Theorem 4.4 If \( \lim_{k \to \infty} \alpha^k = \bar{\alpha} \), then from Theorem 3.2, \( \bar{\alpha} \) is an optimal solution. Furthermore, after \( k \) is large enough, only elements in

\[
\{ t \mid -y_t \nabla f(\bar{\alpha})_t = \max(\max_{\alpha_t < C, y_t = 1} -\nabla f(\bar{\alpha})_t, \max_{\alpha_t > 0, y_t = -1} \nabla f(\bar{\alpha})_t) \}
\]

(4.1)
can still be possibly modified.

Therefore, we tend to guess that if a variable \( \alpha_i \) is equal to \( C \) for several iterations, then at the final solution, it is still at the upper bound. Hence instead of solving the whole problem (2.2), the decomposition method works on a smaller problem:

\[
\begin{align*}
\min_{\alpha_A} & \quad \frac{1}{2} \alpha_A^T Q_{AA} \alpha_A - (p_A - Q_{AN} \alpha_N^k)^T \alpha_A \\
\text{subject to} & \quad 0 \leq (\alpha_A)_t \leq C, t = 1, \ldots, q,
\end{align*}
\]

(4.2)

where \( N = \{1, \ldots, l\} \setminus A \).

Of course this heuristic may fail if the optimal solution of (4.2) is not the corresponding part of that of (2.2). When that happens, the whole problem (2.2) is reoptimized starting from a point \( \alpha \) where \( \alpha_B \) is an optimal solution of (4.2) and \( \alpha_N \) are bounded variables identified before the shrinking process. Note that while solving the shrinked problem (4.2), we only know the gradient \( Q_{AA} \alpha_A + Q_{AN} \alpha_N + p_A \) of (4.2). Hence when problem (2.2) is reoptimized we also have to reconstruct the whole gradient \( \nabla f(\alpha) \), which is quite expensive.

Many implementations began the shrinking procedure near the end of the iterative process, in LIBSVM however, we start the shrinking process from the beginning. The procedure is as follows:

1. After every \( \min(l, 1000) \) iterations, we try to shrink some variables. Note that during the iterative process

\[
\begin{align*}
\min(\{ \nabla f(\alpha^k)_t \mid y_t = -1, \alpha_t < C \}, \{ -\nabla f(\alpha^k)_t \mid y_t = 1, \alpha_t > 0 \}) &= -g_j \\
&< g_i = \max(\{ -\nabla f(\alpha^k)_t \mid y_t = 1, \alpha_t < C \}, \{ \nabla f(\alpha^k)_t \mid y_t = -1, \alpha_t > 0 \})
\end{align*}
\]

(4.3)
as (3.12) is not satisfied yet.

We conjecture that for those

\[
\begin{align*}
g_t &= \begin{cases} -\nabla f(\alpha)_t & \text{if } y_t = 1, \alpha_t < C, \\ \nabla f(\alpha)_t & \text{if } y_t = -1, \alpha_t > 0, \end{cases}
\end{align*}
\]

(4.4)
if
\[ g_t \leq -g_j, \] (4.5)
and \( \alpha_t \) resides at a bound, then the value of \( \alpha_t \) may not change any more. Hence we inactivate this variable. Similarly, for those
\[ g_t \equiv \begin{cases} -\nabla f(\alpha)_t & \text{if } y_t = -1, \alpha_t < C, \\ \nabla f(\alpha)_t & \text{if } y_t = 1, \alpha_t > 0, \end{cases} \] (4.6)
if
\[ -g_t \geq g_i, \] (4.7)
and \( \alpha_t \) is at a bound, it is inactivated. Thus the set \( A \) of activated variables is dynamically reduced in every \( \min(l, 1000) \) iterations.

2. Of course the above shrinking strategy may be too aggressive. Since the decomposition method has a very slow convergence and a large portion of iterations are spent for achieving the final digit of the required accuracy, we would not like those iterations are wasted because of a wrongly shrinked problem (4.2). Hence when the decomposition method first achieves the tolerance
\[ g_i \leq -g_j + 10\epsilon, \]
where \( \epsilon \) is the specified stopping criteria, we reconstruct the whole gradient. Then based on the correct information, we use criteria like (4.4) and (4.6) to inactivate some variables and the decomposition method continues.

Therefore, in \textbf{LIBSVM}, the size of the set \( A \) of (4.2) is dynamically reduced. To decrease the cost of reconstructing the gradient \( \nabla f(\alpha) \), during the iterations we always keep
\[ \bar{G}_i = C \sum_{\alpha_j = C} Q_{ij}, i = 1, \ldots, l. \]
Then for the gradient \( \nabla f(\alpha)_i, i \notin A \), we have
\[ \nabla f(\alpha)_i = \sum_{j=1}^{l} Q_{ij} \alpha_j = \bar{G}_i + \sum_{0 < \alpha_j < C} Q_{ij} \alpha_j. \]

For \( \nu \)-SVC and \( \nu \)-SVR, as the stopping condition (3.25) is different from (3.13), the shrinking strategies (4.5) and (4.7) must be modified. From (4.3), now we have to separate two cases: \( y_t = 1 \) and \( y_t = -1 \). For \( y_t = 1 \), (4.3) becomes
\[ \min \{ -\nabla f(\alpha)_t \mid y_t = 1, \alpha_t > 0 \} = -g_j \]
\[ < g_i = \max \{ -\nabla f(\alpha)_t \mid y_t = 1, \alpha_t < C \} \]
so we inactivate those \( \alpha_t \) with

\[-\nabla f(\alpha)_t \leq -g_j \text{ if } y_t = 1 \text{ and } \alpha_t < C,\]

and

\[-\nabla f(\alpha)_t \geq g_i \text{ if } y_t = 1 \text{ and } \alpha_t > 0.\]

The case for \( y_t = -1 \) is similar.

### 4.2 Caching

Another technique for reducing the computational time is caching. Since \( Q \) is fully dense and may not be stored in the computer memory, elements \( Q_{ij} \) are calculated as needed. Then usually a special storage using the idea of a cache is used to store recently used \( Q_{ij} \) \citep{Joachims1998}. Hence the computational cost of later iterations can be reduced.

Theorem \[4.4\] also supports the use of the cache as in final iterations only some columns of the matrix \( Q \) are still needed. Thus if the cache can contain these columns, we can avoid most kernel evaluations in final iterations.

In \[LIBSVM\] we implement a simple least-recent-use strategy for the cache. We dynamically cache only recently used columns of \( Q_{AA} \) of \[4.2\].

### 4.3 Computational Complexity

The discussion in Section \[3.4\] is about the asymptotic global convergence of the decomposition method. In addition, the linear convergence (Theorem \[3.3\]) is a property of the local convergence rate. Here, we discuss the computational complexity.

The main operations are on finding \( Q_{BN}\alpha_N^k + p_B \) of \[3.3\] and the update of \( \nabla f(\alpha^k) \) to \( \nabla f(\alpha^{k+1}) \). Note that \( \nabla f(\alpha) \) is used in the working set selection as well as the stopping condition. They can be considered together as

\[Q_{BN}\alpha_N^k + p_B = \nabla f(\alpha^k) - Q_{BB}\alpha_B^k, \tag{4.8}\]

and

\[\nabla f(\alpha^{k+1}) = \nabla f(\alpha^k) + Q_{:B}(\alpha_B^{k+1} - \alpha_B^k), \tag{4.9}\]

where \( Q_{:B} \) is the sub-matrix of \( Q \) with column indices in \( B \). That is, at the \( k \)th iteration, as we already have \( \nabla f(\alpha^k) \), the right-hand-side of \[4.8\] is used to construct the sub-problem. After the sub-problem is solved, \[4.9\] is employed to have the next \( \nabla f(\alpha^{k+1}) \). As \( B \) has only two elements and solving the sub-problem is easy, the main cost is \( Q_{:B}(\alpha_B^{k+1} - \alpha_B^k) \) of \[4.9\]. The operation itself takes \( O(2l) \) but if \( Q_{:B} \) is not available in the cache and
each kernel evaluation costs $O(n)$, one column of $Q_{i,B}$ already needs $O(ln)$. Therefore, the complexity is:

1. $\#\text{Iterations} \times O(l)$ if most columns of $Q$ are cached during iterations.

2. $\#\text{Iterations} \times O(nl)$ if most columns of $Q$ are cached during iterations and each kernel evaluation is $O(n)$.

Note that if shrinking is incorporated, $l$ will gradually decrease during iterations.

Unfortunately, so far we do not know much about the complexity of the number of iterations. An earlier work is in [Hush and Scovel, 2003]. However, its result applies only to decomposition methods discussed in [Chang et al., 2000] but not LIBSVM or other existing software.

5 Multi-class classification

We use the “one-against-one” approach [Knerr et al., 1990] in which $k(k - 1)/2$ classifiers are constructed and each one trains data from two different classes. The first use of this strategy on SVM was in [Friedman, 1996; Kreßel, 1999]. For training data from the $i$th and the $j$th classes, we solve the following binary classification problem:

$$\min_{w^{ij}, b^{ij}, \xi^{ij}} \frac{1}{2}(w^{ij})^T w^{ij} + C(\sum_t (\xi^{ij}_t))$$

subject to

$$(w^{ij})^T \phi(x_t) + b^{ij} \geq 1 - \xi^{ij}_t, \text{ if } x_t \text{ in the } i\text{th class},$$

$$(w^{ij})^T \phi(x_t) + b^{ij} \leq -1 + \xi^{ij}_t, \text{ if } x_t \text{ in the } j\text{th class},$$

$$\xi^{ij}_t \geq 0.$$

In classification we use a voting strategy: each binary classification is considered to be a voting where votes can be cast for all data points $x$ - in the end point is designated to be in a class with maximum number of votes.

In case that two classes have identical votes, though it may not be a good strategy, now we simply select the one with the smallest index.

There are other methods for multi-class classification. Some reasons why we choose this “1-against-1” approach and detailed comparisons are in [Hsu and Lin, 2002a].

6 Unbalanced Data

For some classification problems, numbers of data in different classes are unbalanced. Hence some researchers (e.g. [Osuna et al., 1997a]) have proposed to use different penalty param-
eters in the SVM formulation: For example, $C$-SVM becomes
\[
\begin{aligned}
\min_{w,b,\xi} & \quad \frac{1}{2} w^T w + C_+ \sum_{y_i=1} \xi_i + C_- \sum_{y_i=-1} \xi_i \\
\text{subject to} & \quad y_i (w^T \phi(x_i) + b) \geq 1 - \xi_i, \\
& \quad \xi_i \geq 0, i = 1, \ldots, l.
\end{aligned}
\]
Its dual is
\[
\begin{aligned}
\min_{\alpha} & \quad \frac{1}{2} \alpha^T Q \alpha - \mathbf{e}^T \alpha \\
\text{subject to} & \quad 0 \leq \alpha_i \leq C_+, \text{ if } y_i = 1 \quad (6.1) \\
& \quad 0 \leq \alpha_i \leq C_-, \text{ if } y_i = -1 \quad (6.2) \\
& \quad y^T \alpha = 0,
\end{aligned}
\]
Note that by replacing $C$ with different $C_i, i = 1, \ldots, l$, most of the analysis earlier are still correct. Now using $C_+$ and $C_-$ is just a special case of it. Therefore, the implementation is almost the same. A main difference is on the solution of the sub-problem \((3.20)\). Now it becomes:
\[
\begin{aligned}
\min_{\alpha_i, \alpha_j} & \quad \frac{1}{2} \begin{bmatrix} \alpha_i & \alpha_j \end{bmatrix} \begin{bmatrix} Q_{ii} & Q_{ij} \\ Q_{ji} & Q_{jj} \end{bmatrix} \begin{bmatrix} \alpha_i \\ \alpha_j \end{bmatrix} + (Q_{i,N} \alpha_N - 1) \alpha_i + (Q_{j,N} \alpha_N - 1) \alpha_j \\
\text{subject to} & \quad y_i \alpha_i + y_j \alpha_j = \Delta - y_N^T \alpha_N, \quad (6.3) \\
& \quad 0 \leq \alpha_i \leq C_i, 0 \leq \alpha_j \leq C_j,
\end{aligned}
\]
where $C_i$ and $C_j$ can be $C_+$ or $C_-$ depending on $y_i$ and $y_j$.

Following the concept of assigning all bounded values explicitly, if $y_i \neq y_j$, we use the following segment of code:
\[
\begin{aligned}
\text{if}(y[i]\neq y[j]) \\
\{ \\
\quad \text{double delta} = (-G[i]-G[j])/(Q_{i}[i]+Q_{j}[j]+2\times Q_{i}[j]); \\
\quad \text{double diff} = \text{alpha}[i] - \text{alpha}[j]; \\
\quad \text{alpha}[i] += \text{delta}; \\
\quad \text{alpha}[j] += \text{delta}; \\
\quad \text{if}(\text{diff} > 0) \\
\quad \{ \\
\quad \quad \text{if}(\text{alpha}[j] < 0) \\
\quad \quad \{ \\
\quad \quad \quad \text{alpha}[j] = 0; \\
\quad \quad \quad \text{alpha}[i] = \text{diff}; \\
\quad \quad \} \\
\quad \} \\
\} \\
\text{else}
\end{aligned}
\]
if(alpha[i] < 0)
{
    alpha[i] = 0;
    alpha[j] = -diff;
}
if(diff > C_i - C_j)
{
    if(alpha[i] > C_i)
    {
        alpha[i] = C_i;
        alpha[j] = C_i - diff;
    }
}
else
{
    if(alpha[j] > C_j)
    {
        alpha[j] = C_j;
        alpha[i] = C_j + diff;
    }
}

In the above code basically we think the feasible region as a rectangle with length $C_i$ and width $C_j$. Then we consider situations where the line segment $\alpha_i - \alpha_j = y_i(\Delta_1 - y_N^T \alpha_N^k)$ goes to the outside of the rectangular region.

7 Model Selection

*LIBSVM* provides a model selection tool using the RBF kernel: cross validation via parallel grid search. Currently, we support only $C$-SVC where two parameters are considered: $C$ and $\gamma$. They can be easily modified for other kernels such as linear and polynomial.

For medium-sized problems, cross validation might be the most reliable way for model selection. First, the training data is separated to several folds. Sequentially a fold is considered as the validation set and the rest are for training. The average of accuracy on predicting the validation sets is the cross validation accuracy.

Our implementation is as follows. Users provide a possible interval of $C$ (or $\gamma$) with the grid space. Then, all grid points of $(C, \gamma)$ are tried to see which one gives the highest cross validation accuracy. Users then use the best parameter to train the whole training set and generate the final model.

For easy implementation, we consider each SVM with parameters $(C, \gamma)$ as an independent problem. As they are different jobs, we can easily solve them in parallel. Currently,
LIBSVM provides a very simple python interface so that jobs are dispatched to a cluster of computers which share the same file system.

Figure 2: Contour plot of heart_scale included in the LIBSVM package

Note that now under the same \((C, \gamma)\), the one-against-one method is used for training multi-class data. Hence, in the final model, all \(k(k-1)/2\) decision functions share the same \((C, \gamma)\).

LIBSVM also outputs the contour plot of cross validation accuracy. An example is in Figure 2.

8 Probability Estimates

Originally support vector classification (regression) predicts only class label (approximate target value) but not probability information. In the following we briefly describe how we extend SVM for probability estimates. More details are in (Wu et al., 2003) for classification and in (Lin and Weng, 2004) for regression.
Given \( k \) classes of data, for any \( \mathbf{x} \), the goal is to estimate

\[
p_i = p(y = i \mid \mathbf{x}), i = 1, \ldots, k.
\]

Following the setting of the one-against-one (i.e., pairwise) approach for multi-class classification, we first estimated pairwise class probabilities

\[
r_{ij} \approx p(y = i \mid y = i \text{ or } j, \mathbf{x})
\]

using an improved implementation (Lin et al., 2003) of (Platt, 2000):

\[
r_{ij} \approx \frac{1}{1 + e^{Af + B}};
\]

where \( A \) and \( B \) are estimated by minimizing the negative log-likelihood function using known training data and their decision values \( \hat{f} \). Labels and decision values are required to be independent so here we conduct five-fold cross-validation to obtain decision values.

Then the second approach in (Wu et al., 2003) is used to obtain \( p_i \) from all these \( r_{ij} \)'s. It solves the following optimization problem:

\[
\min_{\mathbf{p}} \frac{1}{2} \sum_{i=1}^{k} \sum_{j:j \neq i} (r_{ji} p_i - r_{ij} p_j)^2 \quad \text{subject to } \sum_{i=1}^{k} p_i = 1, p_i \geq 0, \forall i.
\]

The objective function comes from the equality

\[
p(y = j \mid y = i \text{ or } j, \mathbf{x}) \cdot p(y = i \mid \mathbf{x}) = p(y = i \mid y = i \text{ or } j, \mathbf{x}) \cdot p(y = j \mid \mathbf{x})
\]

and can be reformulated as

\[
\min_{\mathbf{p}} \frac{1}{2} \mathbf{p}^T \mathbf{Q} \mathbf{p},
\]

where

\[
Q_{ij} = \begin{cases} 
\sum_{s:s \neq i} r_{si}^2 & \text{if } i = j, \\
r_{ji} r_{ij} & \text{if } i \neq j.
\end{cases}
\]

This problem is convex, so the optimality conditions that there is a scalar \( b \) such that

\[
\begin{bmatrix} Q & \mathbf{e}^T \\
\mathbf{e} & 0 \end{bmatrix} \begin{bmatrix} \mathbf{p} \\ b \end{bmatrix} = \begin{bmatrix} \mathbf{0} \\ 1 \end{bmatrix}.
\]

Here \( \mathbf{e} \) is the \( k \times 1 \) vector of all ones, \( \mathbf{0} \) is the \( k \times 1 \) vector of all zeros, and \( b \) is the Lagrangian multiplier of the equality constraint \( \sum_{i=1}^{k} p_i = 1 \). Instead of directly solving the linear system (8.5), we derive a simple iterative method in the following.

As

\[
-\mathbf{p}^T \mathbf{Q} \mathbf{p} = -\mathbf{p}^T \mathbf{Q}(-b \mathbf{Q}^{-1} \mathbf{e}) = b \mathbf{p}^T \mathbf{e} = b,
\]

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the solution $p$ satisfies

$$Q_mp_t + \sum_{j:j\neq t} Q_{tj}p_j - p^TQp = 0, \text{ for any } t. \quad (8.6)$$

Using (8.6), we consider the following algorithm:

**Algorithm 2**

1. Start with some initial $p_i \geq 0, \forall i$ and $\sum_{i=1}^kp_i = 1$.
2. Repeat ($t = 1, \ldots, k, 1, \ldots$)

   $$p_t \leftarrow \frac{1}{Q_{tt}}[-\sum_{j:j\neq t} Q_{tj}p_j + p^TQp] \quad (8.7)$$

   $$\text{normalize } p \quad (8.8)$$

   until (8.5) is satisfied.

This procedure guarantees to find a global optimum of (8.2). Using some tricks, we do not need to recalculate $p^TQp$ in each iteration. Detailed implementation notes are in Appendix C of [Wu et al., 2003]. We consider a relative stopping condition for Algorithm 2:

$$\|Qp - p^TQp\|_1 = \max_t |(Qp)_t - p^TQp| < 0.005/k.$$ 

When $k$ is large, $p$ will be closer to zero, so we decrease the tolerance by a factor of $k$.

Next, we discuss SVR probability inference. For a given set of training data $D = \{(x_i, y_i) | x_i \in \mathbb{R}^n, y_i \in \mathbb{R}, i = 1, \ldots, l\}$, we suppose that the data are collected from the model:

$$y_i = f(x_i) + \delta_i, \quad (8.9)$$

where $f(x)$ is the underlying function and $\delta_i$ are independent and identically distributed random noises. Given a test data $x$, the distribution of $y$ given $x$ and $D$, $P(y | x, D)$, allows one to draw probabilistic inferences about $y$; for example, one can construct a predictive interval $I = I(x)$ such that $y \in I$ with a pre-specified probability. Denoting $\hat{f}$ as the estimated function based on $D$ using SVR, then $\zeta = \zeta(x) \equiv y - \hat{f}(x)$ is the out-of-sample residual (or prediction error), and $y \in I$ is equivalent to $\zeta \in I - \hat{f}(x)$. We propose to model the distribution of $\zeta$ based on a set of out-of-sample residuals $\{\zeta_i\}_{i=1}^l$ using training data $D$. The $\zeta_i$’s are generated by first conducting a $k$-fold cross validation to get $\hat{f}_j$, $j = 1, \ldots, k$, and then setting $\zeta_i \equiv y_i - \hat{f}_j(x_i)$ for $(x_i, y_i)$ in the $j$th fold. It is conceptually clear that the distribution of $\zeta_i$’s may resemble that of the prediction error $\zeta$.

Figure 3 illustrates $\zeta_i$’s from a real data. Basically, a discretized distribution like histogram can be used to model the data; however, it is complex because all $\zeta_i$’s must be
On the contrary, distributions like Gaussian and Laplace, commonly used as noise models, require only location and scale parameters. In Figure 3 we plot the fitted curves using these two families and the histogram of ζᵢ’s. The figure shows that the distribution of ζᵢ’s seems symmetric about zero and that both Gaussian and Laplace reasonably capture the shape of ζᵢ’s. Thus, we propose to model ζᵢ by zero-mean Gaussian and Laplace, or equivalently, model the conditional distribution of y given ŷ(x) by Gaussian and Laplace with mean ŷ(x).

[Lin and Weng 2004] discussed a method to judge whether a Laplace and Gaussian distribution should be used. Moreover, they experimentally show that in all cases they have tried, Laplace is better. Thus, here we consider the zero-mean Laplace with a density function:

\[ p(z) = \frac{1}{2\sigma} e^{-|z|/\sigma}. \quad (8.10) \]

Assuming that ζᵢ are independent, we can estimate the scale parameter by maximizing the likelihood. For Laplace, the maximum likelihood estimate is

\[ \sigma = \frac{\sum_{i=1}^{l} |\zeta_i|}{l}. \quad (8.11) \]

[Lin and Weng, 2004] pointed out that some “very extreme” ζᵢ may cause inaccurate estimation of σ. Thus, they propose to estimate the scale parameter by discarding ζᵢ’s which exceed ±5 × (standard deviation of ζᵢ). Thus, for any new data x, we consider that

\[ y = \hat{f}(x) + z, \]

where z is a random variable following the Laplace distribution with parameter σ.

In theory, the distribution of ζ may depend on the input x, but here we assume that it is free of x. This is similar to the model (8.1) for classification. Such an assumption works well in practice and leads to a simple model.

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References

Figure 3: Histogram of $\zeta_i$’s from a data set and the modeling via Laplace and Gaussian distributions. The x-axis is $\zeta_i$ using five-fold CV and the y-axis is the normalized number of data in each bin of width 1.


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