

# “Pre-conditioning” for feature selection and regression in high-dimensional problems

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## Abstract

We consider regression problems where the number of predictors greatly exceeds the number of observations. We propose a method for

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variable selection that first estimates the regression function, yielding a “pre-conditioned” response variable. The primary method used for this initial regression is supervised principal components. Then we apply a standard procedure such as forward stepwise selection or the LASSO to the pre-conditioned response variable. In a number of simulated and real data examples, this two-step procedure outperforms forward stepwise selection or the usual LASSO (applied directly to the raw outcome). We also show that under a certain Gaussian latent variable model, application of the LASSO to the pre-conditioned response variable is consistent as the number of predictors and observations increases. Moreover, when the observational noise is rather large, the suggested procedure can give a more accurate estimate than LASSO. We illustrate our method on some real problems, including survival analysis with microarray data.

## 1 Introduction

In this paper we consider the problem of fitting linear (and other related) models to data for which the number of features  $p$  greatly exceeds the number of samples  $n$ . This problem occurs frequently in genomics, for example in microarray studies in which  $p$  genes are measured on  $n$  biological samples.

The problem of model selection for data where number of variables is

typically comparable or much larger than the sample size has received a lot of attention recently. In particular, various penalized regression methods are being widely used as means of selecting the variables having nonzero contribution in a regression model. Among these tools the  $L_1$  penalized regression or LASSO (Tibshirani (1996)) is one of the most popular techniques. The Least Angle Regression (LAR) procedure Efron et al. (2004) provides a method for fast computation of LASSO solution in regression problems. Osborne et al. (2000) derived the optimality conditions associated with the LASSO solution. Donoho & Elad (2003) and Donoho (2004) proved some analytical properties of the  $L_1$  penalization approach for determining the sparsest solution for an under-determined linear system. Some statistical properties of the LASSO-based estimator of the regression parameter have been derived by Knight & Fu (2000). In the context of high-dimensional graphs, Meinshausen & Bühlmann (2006) showed that the variable selection method based on lasso can be consistent if the underlying model satisfies some conditions. Various other model selection criteria have been proposed in high dimensional regression problems. Fan & Li (2005) and Shen & Ye (2002) gave surveys of some of these methods.

However, when the number of variables ( $p$ ) is much larger than the number of observations (precisely  $p_n \sim cn^\xi$  for some  $\xi \in (0, 1)$ ) Meinshausen (2005) showed that the convergence rate of risk of the LASSO estimator can

be quite slow. For finite-dimensional problems, Zou (2005) found a necessary condition for the covariance matrix of the observations, without which the LASSO variable selection approach is inconsistent. Zhao & Yu (2006) derived a related result for the  $p > N$  case.

Various modifications to LASSO have been proposed to ensure that on one hand, the variable selection process is consistent and on the other, the estimated regression parameter has a fast rate of convergence. Fan & Li (2005) proposed the Smoothly Clipped Absolute Deviation (SCAD) penalty for variable selection. ? discussed the asymptotic behavior of this and other related penalized likelihood procedures when the dimensionality of the parameter is growing. Zou (2005) proposed a non-negative Garrote-type penalty (that is re-weighted by the least squares estimate of the regression parameter) and showed that this estimator has adaptivity properties when  $p$  is fixed. Meinshausen (2005) proposed a relaxation to the LASSO penalty after initial model selection to address the problem of high bias of LASSO estimate when  $p$  is very large.

All of these methods try to solve two problems at once: 1) find a good predictor  $\hat{y}$  and 2) find a (hopefully small) subset of variables to form the basis for this prediction. When  $p \gg n$ , these problems are especially difficult. In this paper we suggest that they should be solved separately, rather than both at once. Moreover, the method we propose utilizes the correlation structure

of the predictors, unlike most of the methods cited. We propose a two-stage approach :

- (a) find a consistent predictor  $\hat{y}$  of the true response,
- (b) using the *pre-conditioned* outcome  $\hat{y}$ , apply a model fitting procedure (such as forward stagewise selection or the LASSO) to the data  $(\mathbf{x}, \hat{y})$ .

In this paper we show that the use of  $\hat{y}$  in place of  $y$  in the model selection step (b) can mitigate the effects of noisy features on the selection process under the setting of a *latent variable model* for the response, when the number of predictor variables that are associated with the response grows at a slower rate than the number of observations, even though the nominal dimension of the predictors can grow at a much faster rate.

This paper is organized as follows. In section 2 we define the pre-conditioning method and give an example from a latent variable model. Section 3 discusses a real example from a kidney cancer microarray study, and application of the idea to other settings such as survival analysis. In section 4 we give details of the latent variable model, and show that the LASSO applied to the pre-conditioned response yields a consistent set of predictors, as the number of features and samples goes to infinity. Finally in section 5 we discuss and illustrate the pre-conditioning idea for classification problems.

## 2 Pre-conditioning

Suppose that the feature measurements are  $x_i = (x_{i1}, x_{i2}, \dots, x_{ip})$  and outcome values  $y_i$ , for  $i = 1, 2, \dots, n$ . Our basic model has the form

$$\mathbb{E}(y_i|x_i) = \theta_0 + \sum_{j=1}^p x_{ij}\theta_j, \quad i = 1, 2, \dots, n \quad (1)$$

Two popular methods for fitting this model are forward stepwise selection (FS) and the LASSO Tibshirani (1996). The first method successively enters the variable that most reduces the residual sum of squares, while the second minimizes the penalized criterion

$$J(\theta, \mu) = \sum_i (y_i - \theta_0 + \sum_{j=1}^p \theta_j x_{ij})^2 + \mu \sum_{j=1}^p |\theta_j|. \quad (2)$$

Efron et al. (2004) develop the least angle regression (LAR) algorithm, for fast computation of the LASSO for all values of the tuning parameter  $\mu \geq 0$ .

Usually model selection in the general model (1) is quite difficult when  $p \gg n$ , and our simulations confirm this. To get better results we may need further assumptions about the underlying model relating  $y_i$  to  $x_i$ . In this paper, we assume that  $y_i$  and  $x_i$  are connected via a low-dimensional *latent variable model*, and use a method that we shall refer to as *pre-conditioning* to carry out model selection. In this approach, we first find a consistent estimate  $\hat{y}_i$  by utilizing the latent variable structure, and then apply a fitting procedure such as forward stepwise regression or the LASSO to the

data  $(x_i, \hat{y}_i), i = 1, 2, \dots, n$ . The main technique that we consider for the initial pre-conditioning step is supervised principal components (SPC) (Bair & Tibshirani (2004), Bair et al. (2006)). This method works as follows:

- a) we select the features whose individual correlation with the outcome is large,
- b) using just these features, we compute the principal components of the matrix of features, giving  $\hat{V}_1, \hat{V}_2, \dots, \hat{V}_{\min\{N,p\}}$ . The prediction  $\hat{y}_i$  is the least squares regression of  $y_i$  on the first  $K$  of these components.

Typically we use just the first or first few supervised principal components. Bair et al. (2006) show that under an assumption about the sparsity of the population principal components, as  $p, n \rightarrow \infty$ , supervised principal components gives consistent estimates for the regression coefficients while the usual principal components regression does not. We give details of this model in section 4, and provide a simple example next.

## 2.1 Example: latent variable model

The following example shows the main idea in this paper. Consider a model of the form:

$$Y = \beta_0 + \beta_1 V + \sigma_1 Z \tag{3}$$

In addition, we have measurements on a set of features  $X_j$  indexed by  $j \in \mathcal{A}$ , for which

$$X_j = \alpha_{0j} + \alpha_{1j}V + \sigma_0 e_j, \quad j \in 1, \dots, p. \quad (4)$$

The quantity  $V$  is an unobserved or *latent* variable. The set  $\mathcal{A}$  represents the important features (meaning that  $\alpha_{1j} \neq 0$ , for  $j \in \mathcal{A}$ ) for predicting  $Y_i$ . The errors  $Z_i$  and  $e_{ij}$  are assumed to have mean zero and are independent of all other random variables in their respective models. All random variables  $(V, Z, e_j)$  have a standard Gaussian distribution.

## 2.2 Example 1

For illustration, we generated data on  $p = 500$  features and  $n = 20$  samples, according to this model, with  $\beta_1 = 2$ ,  $\beta_0 = 0$ ,  $\alpha_{0j} = 0$ ,  $\alpha_{1j} = 1$ ,  $\sigma_1 = 2.5$ ,  $\mathcal{A} = \{1, 2, \dots, 20\}$ . Our goal is to predict  $Y$  from  $X_1, X_2, \dots, X_p$ , and in the process, discover the fact that only the first 20 features are relevant. This is a difficult problem. However if we guess (correctly) that the data were generated from model (4), our task is made easier. The left panel of Figure 1 shows the correlations  $\text{Corr}(V, X_j)$  plotted versus  $\text{Corr}(Y, X_j)$  for each feature  $j$ . The first 20 features are plotted in red, and can be distinguished much more easily on the basis of  $\text{Corr}(V, X_j)$  than  $\text{Corr}(Y, X_j)$ . However this requires knowledge of the underlying latent factor  $V$ , which is



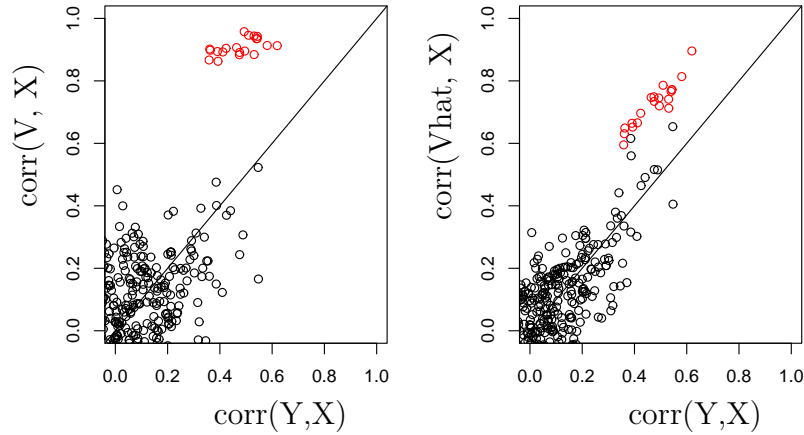


Figure 1: Results for simulated data. Left panel shows the correlation between the true latent variable  $V$  and gene expression  $X$  for each of the genes plotted against the correlation between  $Y$  and gene expression. The truly non-null genes are shown in red. The right panel is the same, except that the estimated latent variable  $\hat{V}$  (from supervised principal components) replaces  $V$ . We see that correlation with either the true or estimated latent factor does a better job at isolating the truly non-null genes.

not observed.

The right panel shows the result when we instead estimate  $V_i$  from the data, using the first supervised principal component. We see that the correlations of each feature with the estimated latent factor also distinguishes the relevant from the irrelevant features.

Not surprisingly, this increased correlation leads to improvements in the

performance of selection methods, as shown in Table 1. We applied four selection methods to the 20 simulated data sets from this model: FS: simple forward stepwise regression; SPC/FS: forward stepwise regression applied to the pre-conditioned outcome from supervised principal components; LASSO, and SPC/LASSO: LASSO applied to pre-conditioned outcome from supervised principal components. The table shows the average number of good variables selected among the first 1,2,5,10, and 20 variables selected, and the corresponding test errors. Pre-conditioning clearly helps both forward selection and the lasso.

### 2.3 Example 2.

The second example was suggested by a referee. It is somewhat artificial but exposes an important assumption that is made by our procedure. We define random variables  $(Y, X_1, X_2, X_3)$  having a Gaussian distribution with mean zero and inverse covariance matrix

$$\Sigma^{-1} = \begin{pmatrix} 2 & 1 & 1 & 1 \\ 1 & 2 & 0 & 1 \\ 1 & 0 & 2 & 1 \\ 1 & 1 & 1 & 2 \end{pmatrix}.$$

We define 297 additional predictors that are  $N(0, 1)$ . The population regression coefficient is  $\beta = (-1, -1, -1, 0, 0, \dots)$  while the (marginal) correlation

Method	Mean # of good variables,				Test error when selecting first:			
	when selecting first:							
	1	5	10	20	1	5	10	20
FS	0.82	0.98	1.12	1.58	267.36	335.4	353.52	357.07
SPC/FS	0.94	2.66	2.86	3.12	241.88	229.47	231.52	232.28
LASSO	0.88	2.05	3.17	3.29	206.54	184.56	186.71	205.85
SPC/LASSO	0.92	4.21	7.75	9.71	212.23	197.07	183.04	178.19

Table 1: *Four selection methods to the 20 simulated data sets from the model of Example 1. Shown are the number of good variables selected among the first 1,2,10, and 20 variables selected, and the corresponding test errors. Preconditioning clearly helps in both cases, and the lasso outperforms forward selection.*

— Method	Mean # of good variables.			
	when selecting first:			
	1	2	3	4
LASSO	1.0	2.0	3.0	3.0
SPC/LASSO	1.0	2.0	2.0	2.0

Table 2: *Performance of LASSO and pre-conditioned LASSO in the second simulation example.*

of each predictor with  $Y$  is  $\rho = (-0.5, -0.5, 0, 0, 0, \dots)$ . Hence  $X_3$  has zero marginal correlation with  $Y$  but has a non-zero partial correlation with  $Y$ , (since  $(\Sigma^{-1})_{14} = 1$ ). The number of good variables when selecting the first 1,2,3 or 4 predictors is shown in Table 2.

We see that the LASSO enters the 3 good predictors first in every simulation, while the pre-conditioned version ignores the 3rd predictor. Supervised principal components screens out this predictor, because it is marginally independent of  $Y$ .

Pre-conditioning with supervised principal components assumes that any important predictor (in the sense of having significantly large nonzero regression coefficient) will also have a substantial marginal correlation with the outcome. This need not be true in practice, but we believe it will often be a good working hypothesis in many practical problems.

— Method	Mean # of good variables.			
	when selecting first:			
	5	10	20	50
LASSO	2.92	5.88	9.04	9.16
SPC/LASSO	2.49	5.13	10.32	19.73

Table 3: *Performance of LASSO and pre-conditioned LASSO in the third simulation example.*

### 2.4 Example 3.

Our third simulation study compares the lasso to the pre-conditioned lasso, in a more neutral setting. We generated 1000 predictors, each having a  $N(0, 1)$  distribution marginally. The first 40 predictors had a pairwise correlation of 0.5, while the remainder were uncorrelated.

The outcome was generated as

$$Y = \sum_{j=1}^{40} \beta_j X_j + \sigma Z \quad (5)$$

with  $Z, \beta_j \sim N(0, 1)$  and  $\sigma = 5$ . Hence the outcome is only a function of the first 40 (“good”) predictors.

We generated 100 datasets from this model: the average number of good variables selected by the lasso and pre-conditioned lasso is shown in Table 3. Note that with just  $n = 50$  samples, the maximum number of predictors

in the model is also 50. While neither method is successful at isolating the bulk of the 40 good predictors, the pre-conditioned lasso finds twice as many good predictors as the lasso in the full model.

## 3 Examples

### 3.1 Kidney cancer data

Zhao et al. (2005) collected gene expression data on 14,814 genes from 177 kidney patients. Survival times (possibly censored) were also measured for each patient, as well as a number of clinical predictors including the grade of the tumor: 1 (good) to 4 (poor).

The data were split into 88 samples to form the training set and the remaining 89 formed the test set. For illustration, in this section we try to predict grade from gene expression. In the next section we predict survival time (the primary outcome of interest) from gene expression. Figure 2 shows the training and test set correlations between grade and its prediction from different methods. We see that for both forward selection and the LASSO, use of the supervised principal component prediction  $\hat{y}$  as the outcome variable (instead of  $y$  itself) makes the procedure less greedy in the training set and yields higher correlations in the test set. While the correlations in the test set are not spectacularly high, for SPC/FS and SPC/LASSO they do result

in a better predictions in the test set.

## 3.2 Application to other regression settings

Extension of our proposal to other kinds of regression outcomes is very simple. The only change is in step (a) of supervised principal components algorithm, where we replace the correlation by an appropriate measure of association. In particular, the likelihood score statistic is an attractive choice.

## 3.3 Survival analysis

Perhaps the most common version of the  $p > n$  regression problem in genomic studies is survival analysis, where the outcome is patient survival (possibly censored). Then we use the partial likelihood score statistic from Cox's proportional hazards score statistic (see Chapter 4 of Kalbfleisch & Prentice (1980)), in step (a) of supervised principal components. After that, we can (conveniently) use the usual least squares version of FS or LASSO in step (2) of the modeling process. Hence the computational advantages of the least angle regression algorithm can be exploited.

Figure 3 shows the result of applying forward stepwise Cox regression (top left panel), forward stepwise selection applied to the SPC predictor (top right panel), LASSO for the Cox model (bottom left panel) and LASSO applied to the SPC predictor (bottom right panel). The bottom left panel was computed

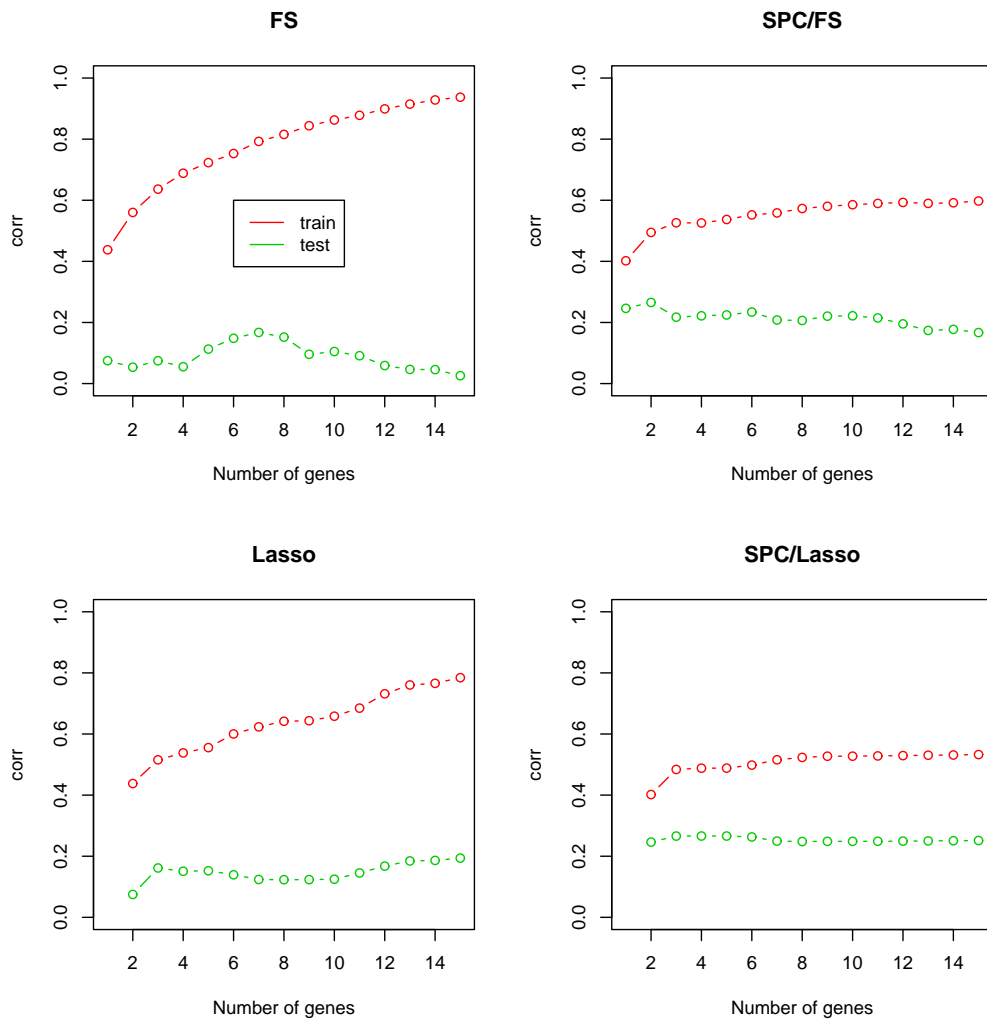


Figure 2: *Kidney cancer data: predicting tumor grade. Correlation of different predictors with the true outcome, in the training and test sets, as more and more genes are entered.*



using the `glm` R package of Park & Hastie (2006), available in the CRAN collection. In each case we obtain a predictor  $\hat{y}$ , and then use  $\hat{y}$  as a covariate in a Cox model, in either the training or test set. The resulting p-values from these Cox models are shown in the figure. We see that forward stepwise Cox regression tends to overfit in the training set, and hence the resulting test-set p-values are not significant. The two stage SPC/FS procedure fits more slowly in the training set, and hence achieves smaller p-values in the test set. “SPC/LASSO” , the LASSO applied to the pre-conditioned response from supervised principal components, performs best and is also computationally convenient: it uses the fast LAR algorithm for the lasso, applied to the pre-conditioned response variable.

The horizontal green line shows the test set p-value of the supervised principal component predictor. We see that the first 10 or 15 genes chosen by the LASSO have captured the signal in this predictor.

We have used the pre-conditioning procedure in real microarray studies. We have found that it is useful to report to investigators not just the best 10 or 15 gene model, but also any genes that have high correlation with this set. The enlarged set can be useful in understanding the underlying biology in experiment, and also for building assays for future clinical use. A given gene might not be well measured on a microarray for a variety of reasons, and hence it is useful to identify surrogate genes that may be used in its place.

Figure 4 shows the average absolute Cox score of the first  $k$  features entered by forward stepwise selection (red) and the pre-conditioned version (green), as  $k$  runs from 1 to 30. The right panel shows the average absolute pairwise correlation of the genes for both methods. We see that the methods enter features of about the same strength, but pre-conditioning enters genes that are more highly correlated with one another.

## 4 Asymptotic analysis

In this section we lay down a mathematical formulation of the problem and pre-conditioning procedure in the context of a latent factor model for the response. We show that the procedure combining SPC with LASSO, under some assumptions about the correlation structure among the variables, leads to asymptotically consistent variable selection in the Gaussian linear model setting. We consider the class of problems where one observes  $n$  independent samples  $(y_i, \mathbf{x}_i)$  where  $y_i$  is a one dimensional response and  $\mathbf{x}_i$  is a  $p$ -dimensional predictor. Individual coordinates of the vector  $\mathbf{x}_i$  are denoted by  $x_{ij}$  where the index  $j \in \{1, \dots, p\}$  correspond to the  $j$ -th predictor. We denote the  $n \times p$  matrix  $((x_{ij}))_{1 \leq i \leq n, 1 \leq j \leq p}$  by  $\mathbf{X}$  and the vector  $(y_i)_{i=1}^n$  by  $Y$ . Henceforth, unless otherwise stated, we do not make a distinction between the realized value  $(Y, \mathbf{X})$  and the random elements (namely, the response and the  $p$  predictors) that they represent.

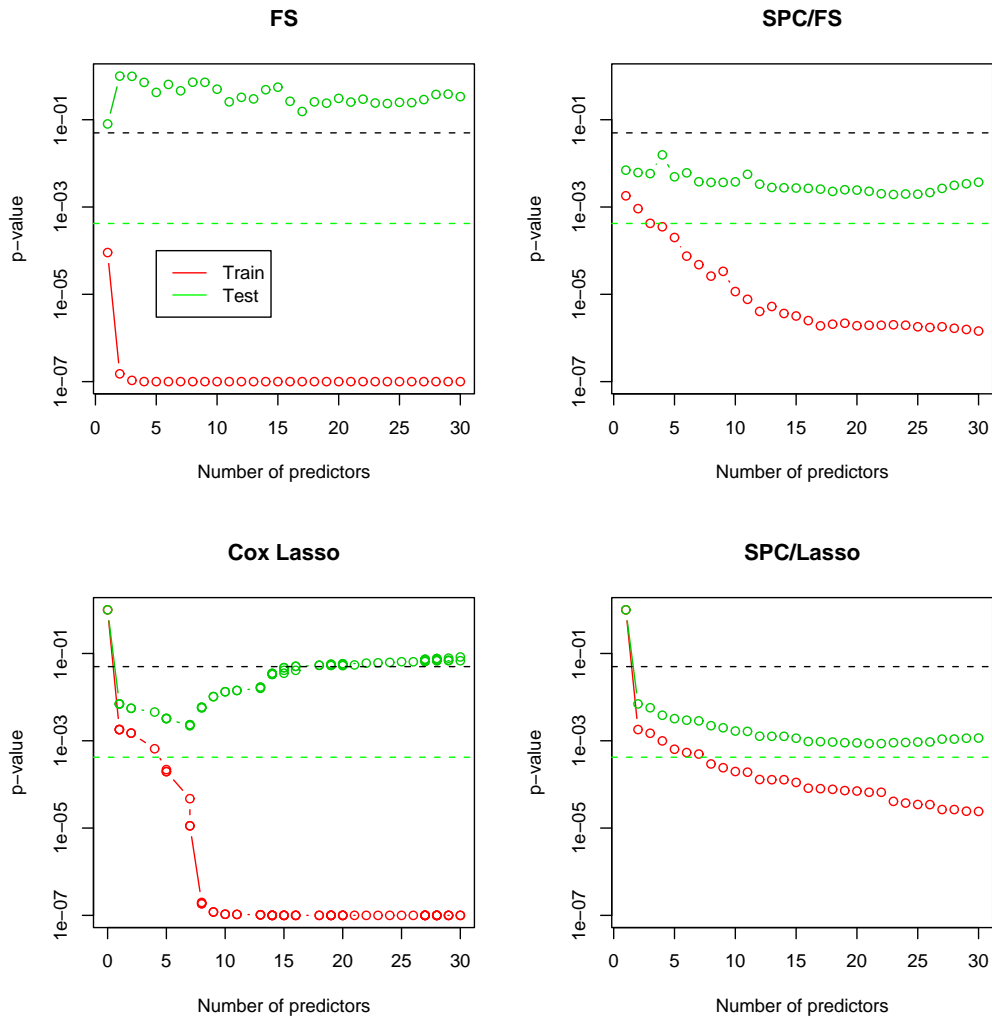


Figure 3: *Kidney cancer data: predicting survival time. Training set p-values (red) and test set p-values (green) for four different selection methods as more and more genes are entered. Horizontal broken lines are drawn at 0.05 (black) and the test set p-value for the supervised principal component predictor 0.00042 (green).*

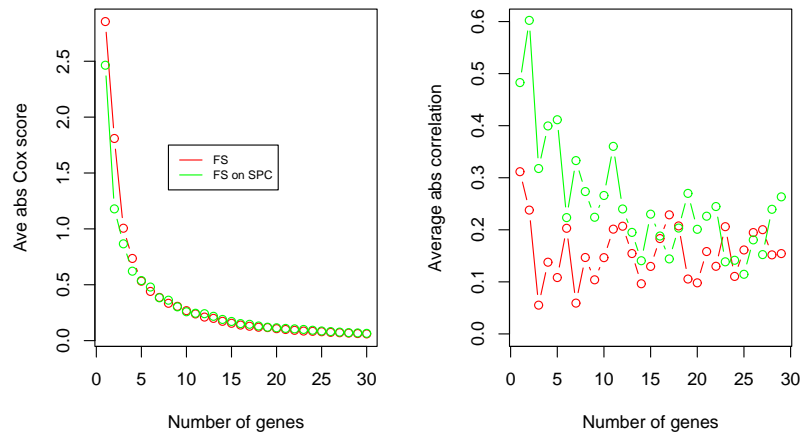


Figure 4: *Kidney cancer data: predicting survival time. Left panel shows the average absolute Cox score of the first  $k$  genes entered by forward stepwise selection (red) and the pre-conditioned version (green), as  $k$  runs from 1 to 30. The right panel shows the average absolute pairwise correlation of the genes for both methods.*

The interest is in identifying the set of predictors  $X_j$  which are (linearly) related to  $Y$ . A regression model will be of the form  $\mathbb{E}(Y|\mathbf{x}) = \theta^T \mathbf{x}$  for some  $\theta \in \mathbb{R}^p$ . Here we assume that the joint distribution of  $\mathbf{X}$  is Gaussian with zero mean and covariance matrix  $\Sigma \equiv \Sigma_p$ . The relationship between  $Y$  and  $\mathbf{X}$  is assumed to be specified by a latent component model to be described below.

## 4.1 Model for $\mathbf{X}$

Suppose that the spectral decomposition of  $\Sigma$  is given by  $\Sigma = \sum_{k=1}^p \ell_k \mathbf{u}_k \mathbf{u}_k^T$ , where  $\ell_1 \geq \dots \geq \ell_p \geq 0$  and  $\mathbf{u}_1, \dots, \mathbf{u}_p$  form an orthonormal basis of  $\mathbb{R}^p$ . We consider the following model for  $\Sigma$ .

Assume that there exists an  $M \geq 1$  such that

$$\ell_k = \lambda_k + \sigma_0^2, \quad k = 1, \dots, M, \quad \text{and} \quad \ell_k = \sigma_0^2, \quad k = M + 1, \dots, p, \quad (6)$$

where  $\lambda_1 \geq \dots \geq \lambda_M > 0$  and  $\sigma_0 > 0$ . This model will be referred to as the “noisy factor model”. To see this, notice that under the Gaussian assumption the matrix  $\mathbf{X}$  can be expressed as

$$\mathbf{X} = \sum_{k=1}^M \sqrt{\lambda_k} \mathbf{v}_k \mathbf{u}_k^T + \sigma_0 \mathbf{E} \quad (7)$$

where  $\mathbf{v}_1, \dots, \mathbf{v}_M$  are i.i.d.  $N_n(0, I)$  vectors (the factors), and  $\mathbf{E}$  is an  $n \times p$  matrix with i.i.d.  $N(0, 1)$  entries, and is independent of  $\mathbf{v}_1, \dots, \mathbf{v}_M$ . This matrix is viewed as a noise matrix.

In the analysis presented in this paper throughout we use (7) as the model for  $\mathbf{X}$ , even though it can be shown that the analysis applies even in the case where  $\ell_{K+1}, \dots, \ell_p$  are decreasing and sufficiently well separated from  $\ell_1, \dots, \ell_K$ .

## 4.2 Model for $Y$

Assume the following regression model for  $Y$ . Note that this is a more general version of (3), even though we assume that  $Y$  has (unconditional) mean 0.

$$Y = \sum_{k=1}^K \beta_k \mathbf{v}_k + \sigma_1 Z, \quad (8)$$

where  $\sigma_1 > 0$ ,  $1 \leq K \leq M$ , and  $Z$  has  $N_n(0, I)$  distribution and is independent of  $\mathbf{X}$ .

## 4.3 Least squares and feature selection

We derive expressions for the marginal correlations between  $Y$  and  $X_j$ , for  $j = 1, \dots, p$  and the (population) least squares solution, viz.  $\theta := \arg \min_{\zeta} \mathbb{E} \| Y - \mathbf{X}\zeta \|_2^2$ , in terms of the model parameters. Let  $\mathcal{P} := \{1, \dots, p\}$ . The marginal correlation between  $\mathbf{X} = (X_j)_{j=1}^p$  and  $Y$  is given by

$$\Sigma_{\mathcal{P}y} := (\mathbb{E}(X_j Y))_{j=1}^p = \sum_{k=1}^K \beta_k \sqrt{\lambda_k} \mathbf{u}_k. \quad (9)$$

The population regression coefficient of  $Y$  on  $\mathbf{X}$ , is given by

$$\begin{aligned}
\theta = \Sigma^{-1}\Sigma_{\mathcal{P}y} &= \left[ \sum_{k=1}^M \lambda_k \mathbf{u}_k \mathbf{u}_k^T + \sigma_0^2 I \right]^{-1} \left[ \sum_{k=1}^K \beta_k \sqrt{\lambda_k} \mathbf{u}_k \right] \\
&= \left[ \sum_{k=1}^M \frac{1}{\lambda_k + \sigma_0^2} \mathbf{u}_k \mathbf{u}_k^T + \frac{1}{\sigma_0^2} \left( I - \sum_{k=1}^M \mathbf{u}_k \mathbf{u}_k^T \right) \right] \left[ \sum_{k=1}^K \beta_k \sqrt{\lambda_k} \mathbf{u}_k \right] \\
&= \sum_{k=1}^K \beta_k \frac{\sqrt{\lambda_k}}{\lambda_k + \sigma_0^2} \mathbf{u}_k = \sum_{k=1}^K \beta_k \ell_k^{-1} \sqrt{\lambda_k} \mathbf{u}_k. \tag{10}
\end{aligned}$$

Now, define  $\mathbf{w}_j = (\sqrt{\lambda_1} u_{j1}, \dots, \sqrt{\lambda_K} u_{jK})^T$ . Let  $\mathcal{D} = \{j : \|\mathbf{w}_j\|_2 \neq 0\}$ . Observe that  $\Sigma_{jy} = \beta^T \mathbf{w}_j$ , and  $\theta_j = \beta^T D_K^{-1} \mathbf{w}_j$ , where  $D_K = \text{diag}(\ell_1, \dots, \ell_K)$ . So if we define  $\mathcal{B} := \{j : \Sigma_{jy} \neq 0\}$ , and  $\mathcal{A} = \{j : \theta_j \neq 0\}$ , then  $\mathcal{B} \subset \mathcal{D}$  and  $\mathcal{A} \subset \mathcal{D}$ .

This gives rise to the regression model:

$$Y = \mathbf{X}\theta + \sigma_\varepsilon \varepsilon, \tag{11}$$

where

$$\sigma_\varepsilon^2 = \sigma_{yy} - \Sigma_{y\mathcal{P}} \Sigma^{-1} \Sigma_{\mathcal{P}y} = \sigma_1^2 + \sum_{k=1}^K \beta_k^2 - \sum_{k=1}^K \beta_k^2 \frac{\lambda_k}{\lambda_k + \sigma_0^2} = \sigma_1^2 + \sigma_0^2 \beta^T D_K^{-1} \beta, \tag{12}$$

and  $\varepsilon$  has i.i.d.  $N(0, 1)$  entries and is independent of  $\mathbf{X}$ .

Note also that, the population partial covariance between  $Y$  and  $\mathbf{X}_C$  given  $\mathbf{X}_D$  (given by  $\Sigma_{yC|D} := \Sigma_{yC} - \Sigma_{yD} \Sigma_{DD}^{-1} \Sigma_{DC}$ ), for any subset  $C \subset \mathcal{D}^c$ , where  $\mathcal{D}^c := \mathcal{P} \setminus \mathcal{D}$ , is 0. However the corresponding statement is not true in general if one replaces  $\mathcal{D}$  by either  $\mathcal{A}$  or  $\mathcal{B}$ . Therefore, ideally, one would

like to identify  $\mathcal{D}$ . However, it may not be possible to accomplish this in general when the dimension  $p$  grows with the sample size  $n$ . Rather, we define the feature selection problem as the problem of identifying  $\mathcal{A}$ , while the estimation problem is to obtain an estimate of  $\theta$  from model (11).

Observe that, if either  $K = 1$  or  $\lambda_1 = \dots = \lambda_K$ , then  $\mathcal{A} = \mathcal{B}$ . In the former case we actually have  $\mathcal{A} = \mathcal{B} = \mathcal{D}$ . In these special cases, the feature selection problem reduces to finding the set  $\mathcal{B}$ , which may be done (under suitable identifiability conditions) just by computing the sample marginal correlations between the response and the predictors and selecting those variables (coordinates) for which the marginal correlation exceeds an appropriate threshold. The major assumptions that we shall make here for solving the problem are that (i)  $\mathcal{A} \subset \mathcal{B}$ , (ii)  $\mathcal{B}$  can be identified from the data (at least asymptotically), (iii) cardinality of  $\mathcal{B}$  (and hence that of  $\mathcal{A}$ ) is small compared to  $n$ , and (iv) the contribution of the coordinates  $\mathcal{B}^c$  in the vectors  $\mathbf{u}_1, \dots, \mathbf{u}_K$  is asymptotically negligible in an  $L^2$  sense. If these conditions are satisfied, then it will allow for the identification of  $\mathcal{A}$ , even as dimension increases with the sample size. We make these (and other) conditions more precise in Section 4.7.



## 4.4 SPC as a preconditioner

The formulation in the previous section indicates that one may use some penalized regression methods to estimate the regression parameter  $\theta$  from the model (11). However, standard methods like LASSO do not use the covariance structure of the data. Therefore if one uses the underlying structure for  $\Sigma$ , and has good estimates of the parameters  $(\mathbf{u}_k, \ell_k)$ , then one can hope to be able to obtain a better estimate  $\theta$ , as well as identify  $\mathcal{A}$  as  $n \rightarrow \infty$ .

We focus on (7) and (8). In general it is not possible to eliminate the contribution of  $\mathbf{E}$  entirely from an estimate of  $\mathbf{v}_k$ , even if we had perfect knowledge of  $(\mathbf{u}_k, \ell_k)$ . To understand this, note that, the conditional distribution of  $\mathbf{v}_k$  given  $\mathbf{X}$  is the same as the conditional distribution of  $\mathbf{v}_k$  given  $\mathbf{X}\mathbf{u}_k$ . The latter distribution is normal with mean  $\frac{\sqrt{\lambda_k}}{\ell_k}\mathbf{X}\mathbf{u}_k$  and covariance matrix  $\frac{\sigma_0^2}{\ell_k}I_n$ . This means that any reasonable procedure that estimates the parameters  $(\mathbf{u}_k, \ell_k)$  can only hope to reduce the effect of the measurement noise in  $Y$ , viz.  $\sigma_1 Z$ .

Keeping these considerations in mind, we employ a two stage procedure described in the following section for estimating  $\theta$ . In order to fit the model (11) using SPC procedure, it is necessary to estimate the eigenvectors  $\mathbf{u}_k$ ,  $k = 1, \dots, M$ . When  $\frac{p}{n}$  is large (in the sense that the fraction does not converge to 0 as  $n \rightarrow \infty$ ), in general it is not possible to estimate  $\mathbf{u}_k$  consistently. However, if  $\mathbf{u}_k$  are sparse, in the sense of having say  $q$  non-zero components,

where  $\frac{q}{n} \rightarrow 0$ , then Bair et al. (2006) showed that under suitable identifiability conditions, it is possible to get asymptotically consistent estimators of  $\mathbf{u}_1, \dots, \mathbf{u}_K$ , where the consistency is measured in terms of convergence of the  $L^2$  distance between the parameter and its estimator.

## 4.5 Algorithm

In this section we present the algorithm in detail.

**Step 1** Estimate  $(\mathbf{u}_1, \ell_1), \dots, (\mathbf{u}_K, \ell_K)$  by SPC procedure in which only those predictors  $X_j$  whose empirical correlation with response  $Y$  is above a threshold  $\tau_n$  are used in the eigen-analysis. Call these estimates  $\{\tilde{\mathbf{u}}_k, \tilde{\ell}_k\}_{k=1}^K$ .

**Step 2** Let  $\tilde{P}_K := Proj(\hat{V}_1, \dots, \hat{V}_K)$  be the projection onto  $\hat{V}_1, \dots, \hat{V}_K$ , where  $\hat{V}_k := \frac{1}{\sqrt{\tilde{\ell}_k}} \mathbf{X} \tilde{\mathbf{u}}_k$  is the  $k$ -th principal component of the predictors (under the SPC procedure). Define  $\tilde{Y} = \tilde{P}_K Y$ .

**Step 3** Estimate  $\theta$  from the linear model  $\tilde{Y} = \mathbf{X}\theta + \text{error}$ , using the LASSO approach with penalty  $\mu_n > 0$ .

Since by definition  $\frac{1}{n} \langle \mathbf{X} \tilde{\mathbf{u}}_k, \mathbf{X} \tilde{\mathbf{u}}_{k'} \rangle = \tilde{\ell}_k \delta_{kk'}$ , it follows that

$$\tilde{P}_K = Proj(\mathbf{X} \tilde{\mathbf{u}}_1, \dots, \mathbf{X} \tilde{\mathbf{u}}_K) = \sum_{k=1}^K \frac{1}{\|\mathbf{X} \tilde{\mathbf{u}}_k\|^2} (\mathbf{X} \tilde{\mathbf{u}}_k) (\mathbf{X} \tilde{\mathbf{u}}_k)^T = \sum_{k=1}^K \frac{1}{\tilde{\ell}_k} \frac{1}{n} (\mathbf{X} \tilde{\mathbf{u}}_k) (\mathbf{X} \tilde{\mathbf{u}}_k)^T. \quad (13)$$

## 4.6 Analysis of the projection

We present an expansion of the projected response  $\tilde{Y} := \tilde{P}_K Y$  that will be useful for all the asymptotic analyses that follow. Using the representation of  $\tilde{P}_K$  in (13) and invoking (7) and (8), we get

$$\begin{aligned}
\tilde{Y} &= \sum_{k=1}^K \frac{\beta_k}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{X}\tilde{\mathbf{u}}_k, \mathbf{v}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k + \sum_{k=1}^K \sum_{k' \neq k}^K \frac{\beta_{k'}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{X}\tilde{\mathbf{u}}_k, \mathbf{v}_{k'} \rangle \mathbf{X}\tilde{\mathbf{u}}_k + \sigma_1 \sum_{k=1}^K \frac{1}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{X}\tilde{\mathbf{u}}_k, Z \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&= \sum_{k=1}^K \frac{\beta_k \sqrt{\lambda_k}}{\tilde{\ell}_k} \frac{1}{n} \|\mathbf{v}_k\|^2 \langle \mathbf{u}_k, \tilde{\mathbf{u}}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k + \sum_{k=1}^K \sum_{l \neq k}^M \frac{\beta_k \sqrt{\lambda_l}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{v}_l, \mathbf{v}_k \rangle \langle \mathbf{u}_l, \tilde{\mathbf{u}}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&\quad + \sum_{k=1}^K \sum_{k' \neq k}^K \sum_{l=1}^M \frac{\beta_{k'} \sqrt{\lambda_l}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{v}_l, \mathbf{v}_{k'} \rangle \langle \mathbf{u}_l, \tilde{\mathbf{u}}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&\quad + \sigma_0 \sum_{k=1}^K \sum_{k'=1}^K \frac{\beta_{k'}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{E}\tilde{\mathbf{u}}_k, \mathbf{v}_{k'} \rangle \mathbf{X}\tilde{\mathbf{u}}_k + \sigma_1 \sum_{k=1}^K \frac{1}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{X}\tilde{\mathbf{u}}_k, Z \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&= \mathbf{X}\theta + \mathbf{X} \sum_{k=1}^K \beta_k \sqrt{\lambda_k} \left( \frac{1}{\tilde{\ell}_k} \frac{\|\mathbf{v}_k\|^2}{n} \langle \mathbf{u}_k, \tilde{\mathbf{u}}_k \rangle \tilde{\mathbf{u}}_k - \frac{1}{\ell_k} \mathbf{u}_k \right) + \sum_{k=1}^K \sum_{k' \neq k}^K \frac{\beta_{k'} \sqrt{\lambda_{k'}}}{\tilde{\ell}_k} \frac{\|\mathbf{v}_{k'}\|^2}{n} \langle \mathbf{u}_{k'}, \tilde{\mathbf{u}}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&\quad + \sum_{k=1}^K \sum_{k' \neq k}^K \frac{\beta_{k'} \sqrt{\lambda_k}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{v}_k, \mathbf{v}_{k'} \rangle \langle \mathbf{u}_k, \tilde{\mathbf{u}}_k \rangle \mathbf{X}\tilde{\mathbf{u}}_k \\
&\quad + \sigma_0 \sum_{k=1}^K \sum_{k'=1}^K \frac{\beta_{k'}}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{E}\tilde{\mathbf{u}}_k, \mathbf{v}_{k'} \rangle \mathbf{X}\tilde{\mathbf{u}}_k + \sigma_1 \sum_{k=1}^K \frac{1}{\tilde{\ell}_k} \frac{1}{n} \langle \mathbf{X}\tilde{\mathbf{u}}_k, Z \rangle \mathbf{X}\tilde{\mathbf{u}}_k + R_n, \tag{14}
\end{aligned}$$

for some vector  $R_n \in \mathbb{R}^n$ . This is an asymptotically unbiased regression model for estimating  $\theta$  provided  $(\tilde{\mathbf{u}}_k, \tilde{\ell}_k)_{k=1}^K$  is an asymptotically consistent estimator for  $(\mathbf{u}_k, \ell_k)_{k=1}^K$ .

## 4.7 Assumptions

In this section we give sufficient conditions for the consistency of the variable selection aspect of the SPC preconditioning procedure. The methods of Zou (2005) and Knight & Fu (2000) are not applicable in our situation since the dimension is growing with the sample size. For most parts, we make assumptions similar to those in Meinshausen & Bühlmann (2006) for the relationship among the variables.

**A1** The eigenvalues  $\lambda_1, \dots, \lambda_M$  satisfy

- (i)  $\lambda_1 > \dots > \lambda_K > \lambda_{K+1} \geq \dots \geq \lambda_M \geq 0$ .
- (ii)  $\min_{1 \leq k \leq K} (\lambda_k - \lambda_{k+1}) \geq C_0$  for some  $C_0 > 0$  (fixed).
- (iii)  $\lambda_1 \leq \Lambda_{\max}$  for some  $\Lambda_{\max}$  fixed. Also,  $\sigma_0$  is fixed.

**A2**  $\sigma_1^2 = O(n^{\kappa_0})$  for some  $\kappa_0 \in (0, \frac{1}{2})$ .

**A3**  $|\mathcal{A}| = q_n$ ,  $|\mathcal{B}| = \bar{q}_n$  such that  $\bar{q}_n = O(n^{\kappa_1})$  for some  $\kappa_1 \in (0, \frac{1}{2})$ .

**A3'**  $p_n$ , the number of variables, satisfies the condition that there is an  $\alpha > 0$  such that  $\log p_n = O(n^\alpha)$  for some  $\alpha \in (0, 1)$ .

**A4** There exists a  $\rho_n$  satisfying  $\rho_n n^{1/2} (\log p_n)^{-1/2} \rightarrow \infty$  as  $n \rightarrow \infty$  such that

$$\min_{j \in \mathcal{B}} \left| \frac{\Sigma_{jy}}{\sqrt{\Sigma_{jj} \sigma_{yy}}} \right| \geq \rho_n. \quad (15)$$

**A5** There exists a  $\bar{\delta}_n$  with  $\bar{\delta}_n = o(\frac{\bar{q}_n}{n \log n})$  such that  $\sum_{j \notin \mathcal{B}} \|\mathbf{w}_j\|_2^2 \leq \bar{\delta}_n$ .

**A6** There exists an  $\eta_n > 0$  satisfying  $\eta_n^{-1} = O(n^{\kappa_2})$  for some  $\kappa_2 < \frac{1}{2}(1 - \kappa_0 \vee \kappa_1)$ , such that

$$\min_{j \in \mathcal{A}} |\theta_j| \geq \eta_n. \quad (16)$$

**A7** There exists a  $\delta \in (0, 1)$  such that

$$\|\Sigma_{\mathcal{A}^c \mathcal{A}} \Sigma_{\mathcal{A} \mathcal{A}}^{-1} \text{sign}(\theta_{\mathcal{A}})\|_{\infty} < \delta. \quad (17)$$

**A8** There is a  $\vartheta < \infty$  such that,

$$\max_{j \in \mathcal{A}} \|\Sigma_{\mathcal{A}_j \mathcal{A}_j}^{-1} \Sigma_{\mathcal{A}_j j}\|_1 < \vartheta, \quad \text{where } \mathcal{A}_j := \mathcal{A} \setminus \{j\}. \quad (18)$$

A few remarks about these conditions are in order. First, condition **A1** about the separation of the eigenvalues is not really necessary, but is assumed to avoid the issue of un-identifiability of an eigenvector. However, the scaling of the eigenvalues is important for the analysis. We remark that it is not necessary that the eigenvalues  $\lambda_1, \dots, \lambda_M$  are the  $M$  largest eigenvalues of  $\Sigma$  in order for the conclusions to hold. All that is necessary is that these are the leading eigenvalues of the matrix  $\Sigma_{\mathcal{D}\mathcal{D}}$ , and there is enough separation from the other eigenvalues of  $\Sigma$ . However, this assumption is made to simplify the exposition.

Next, the condition that  $\bar{q}_n = o(n)$  (implicit from condition **A3**) is necessary for the consistency of the estimated eigenvectors  $\tilde{\mathbf{u}}_k$  from Supervised

PCA. Condition **A4** is necessary for the identifiability of the set  $\mathcal{B}$ . **A5** implies that the contribution of the predictors  $\{X_j : j \in \mathcal{D} \setminus \mathcal{B}\}$  is negligible in our analysis. Note that  $\bar{\delta}_n$  is essentially measuring the “selection bias” for restricting analysis to  $\mathcal{B}$  rather than  $\mathcal{D}$ . Again, the assumption about the rate of decay of  $\bar{\delta}_n$  can be relaxed at the cost of more involved analysis and smaller range of values for  $\mu_n$  (see also the remark following *Corollary 1*). Too large a value of  $\bar{\delta}_n$  may mean that we may not be able to select the variables consistently. Condition **A6** is an identifiability condition for set  $\mathcal{A}$ .

Condition **A7** is needed to guarantee consistency of the variable selection by LASSO after projection. This condition was shown to be necessary for variable selection in finite dimensional LASSO regression by Zou (2005) and also, implicitly by Meinshausen & Bühlmann (2006). Zhao & Yu (2006) termed this the “irrepresentable condition” and showed that it is nearly necessary and sufficient for consistency of model selection by LASSO when  $p, n \rightarrow \infty$ . A sufficient condition for this to hold is that  $\max_{j \in \mathcal{A}^c} \|\Sigma_{\mathcal{A}\mathcal{A}}^{-1} \Sigma_{\mathcal{A}j}\|_1 < \delta$ . Observe that  $\Sigma_{\mathcal{A}\mathcal{A}}^{-1} \Sigma_{\mathcal{A}j}$  is the population regression coefficient in the regression of  $X_j$  on  $\{X_l : l \in \mathcal{A}\}$ . If we are using the estimate  $\widehat{\theta}^{\mathcal{B}, \mu}$  then (see proof of *Lemma 2*) we can replace **A7** by the weaker requirement

$$\|\Sigma_{\mathcal{A}^c \cap \mathcal{B}, \mathcal{A}} \Sigma_{\mathcal{A}\mathcal{A}}^{-1} \text{sign}(\theta_{\mathcal{A}})\|_{\infty} < \delta, \quad \text{for some } \delta \in (0, 1).$$

## 4.8 LASSO solution

We use the symbol  $\mu$  to denote the penalty parameter in LASSO. The LASSO estimate of  $\theta$ , after preconditioning, is given by

$$\widehat{\theta}^\mu = \arg \min_{\zeta \in \mathbb{R}^p} \frac{1}{n} \|\widetilde{Y} - \mathbf{X}\zeta\|_2^2 + \mu \|\zeta\|_1. \quad (19)$$

We also define the *selected LASSO estimate* of  $\theta$  by

$$\widehat{\theta}^{\mathcal{B},\mu} = \arg \min_{\zeta \in \mathbb{R}^p, \zeta_{\mathcal{B}^c} = 0} \frac{1}{n} \|\widetilde{Y} - \mathbf{X}\zeta\|_2^2 + \mu \|\zeta\|_1. \quad (20)$$

For future use, we define the *restricted LASSO estimate* of  $\theta$  to be

$$\widehat{\theta}^{\mathcal{A},\mu} = \arg \min_{\zeta \in \mathbb{R}^p, \zeta_{\mathcal{A}^c} = 0} \frac{1}{n} \|\widetilde{Y} - \mathbf{X}\zeta\|_2^2 + \mu \|\zeta\|_1. \quad (21)$$

The notations used here follow Meinshausen & Bühlmann (2006).

## 4.9 Consistency of variable selection

We shall prove most of our consistency results for the estimate  $\widehat{\theta}^{\mathcal{B},\mu}$  and indicate how (and under what conditions) the same may be proved for the unrestricted estimator  $\widehat{\theta}^\mu$ . As we shall see, when the model assumptions hold the former estimator is more reliable under a wider range of possible dimensions. The latter can consistently select the model essentially when  $p_n = O(n^\kappa)$  for some  $\kappa < \infty$ . In order to prove these results, it will be convenient for us to assume that we have two independent subsamples of

size  $n$  each, so that the total sample size is  $2n$ . And we also assume that **Step 1** of the variable selection algorithm (estimating  $\mathcal{B}$ ) is performed on the first subsample and the other steps are performed on the second subsample. This extra assumption simplifies our proofs (see the proof of *Proposition 4* in the Appendix) somewhat. Further, we shall assume that  $K$ , the number of latent components for response  $Y$ , is known. The results presented here hold uniformly w.r.t. the parameters satisfying assumptions **A1-A8**.

Let  $\widehat{\mathcal{A}}_{\widehat{\mathcal{B}},\mu}$  (resp.  $\widehat{\mathcal{A}}_\mu$ ) denote the set of nonzero coordinates of the vector  $\widehat{\theta}^{\widehat{\mathcal{B}},\mu}$  (resp.  $\widehat{\theta}^\mu$ ). Whenever the context is clear, we shall drop the subscripts from  $\widehat{\mathcal{A}}$ . In the following  $\zeta$  will be used to denote a generic value of the parameter.

**Proposition 1 :** Let  $\widehat{\mathcal{B}}$  denote the set of coordinates selected by the preliminary thresholding scheme of SPC with threshold  $\tau_n$ . Given any  $c_1 > 1$ , and there is a  $\tau_n(c_1) := d_1 \sqrt{\frac{\log pn}{n}}$ , for some constant  $d_1 > 2$ , such that, for  $n \geq n_{c_1}$ ,

$$\mathbb{P}(\widehat{\mathcal{B}} = \mathcal{B}) \geq 1 - n^{-c_1}. \quad (22)$$

*Proposition 1* tells us that we can restrict our analysis to the set  $\mathcal{B}$  while analyzing the effect of preconditioning, and studying the estimator  $\widehat{\theta}^{\widehat{\mathcal{B}},\mu}$ . Our next result is about the behavior of the estimated eigenvalues and eigenvectors of the matrix  $\mathbf{S}_{\widehat{\mathcal{B}}\widehat{\mathcal{B}}} := \frac{1}{n} \mathbf{X}_{\widehat{\mathcal{B}}}^T \mathbf{X}_{\widehat{\mathcal{B}}}$ . This result can be proved along the lines



of *Theorem 3.2* in Paul (2005), (see also Bair et al. (2006)) and is omitted.

**Proposition 2 :** Let  $(\bar{\mathbf{u}}_{\mathcal{B}k}, \bar{\ell}_k)_{k=1}^K$  denote the first  $k$  eigenvector-eigenvalue pairs of  $\Sigma_{\mathcal{B}\mathcal{B}}$ . Suppose that assumptions **A1-A5** hold. Then there are functions  $\gamma_i = \gamma_i(\lambda_1/\sigma_0, \dots, \lambda_M/\sigma_0)$ ,  $i = 1, 2$  such that, given  $c_2 > 0$  there exist  $d_2, d'_2 \geq 1$  so that,

$$\begin{aligned} \mathbb{P}(\max_{1 \leq k \leq K} \|\tilde{\mathbf{u}}_{\mathcal{B}k} - \bar{\mathbf{u}}_{\mathcal{B}k}\|_2 > d_2 \sigma_0 \gamma_1 \sqrt{\frac{\bar{q}_n \vee \log n}{n}} (1 + \sqrt{\frac{\bar{q}_n \log n}{n}}), \widehat{\mathcal{B}} = \mathcal{B}) &= O(n^{-c}), \\ \mathbb{P}(\max_{1 \leq k \leq K} |\tilde{\ell}_k - \bar{\ell}_k| > d'_2 \sigma_0^2 \gamma_2 (\sqrt{\frac{\log n}{n}} + \frac{\bar{q}_n \log n}{n}), \widehat{\mathcal{B}} = \mathcal{B}) &= O(n^{-c}). \end{aligned}$$

**Theorem 1 :** Suppose that assumptions **A1-A8** hold. If  $\mu = \mu_n$  satisfies  $\mu_n = o(n^{-\kappa_2})$  and  $\mu_n n^{\frac{1}{2}(1-\kappa_0 \vee \kappa_1)} \rightarrow \infty$  as  $n \rightarrow \infty$ , then there exists some  $c > 1$  such that, for large enough  $n$ ,

$$\mathbb{P}(\widehat{\mathcal{A}} \subset \mathcal{A}) \geq 1 - O(n^{-c}), \quad (23)$$

where  $\widehat{\mathcal{A}} = \widehat{\mathcal{A}}_{\widehat{\mathcal{B}}, \mu_n}$ . If moreover,  $p_n$  is such that  $\frac{q_n \log p_n}{n} = o(1)$  as  $n \rightarrow \infty$ , then (23) holds with  $\widehat{\mathcal{A}} = \widehat{\mathcal{A}}_{\mu_n}$ .

**Theorem 2 :** With  $\mu = \mu_n$  and  $\widehat{\mathcal{A}}$  as in *Theorem 1*, there exists  $c > 1$  such that,

$$\mathbb{P}(\mathcal{A} \subset \widehat{\mathcal{A}}) \geq 1 - O(n^{-c}). \quad (24)$$

Clearly, *Theorem 1* and *Theorem 2* together imply that the SPC/LASSO procedure asymptotically selects the correct set of predictors under the stated assumptions. The proofs of these critically rely on the following three results.

**Lemma 1 :** Given  $\theta \in \mathbb{R}^p$ , let  $G(\theta)$  be the vectors whose components are defined by

$$G_j(\theta) = -\frac{2}{n} \langle \tilde{Y} - \mathbf{X}\theta, X_j \rangle \quad (25)$$

A vector  $\hat{\theta}$  with  $\hat{\theta}_j = 0$  for all  $j \in \mathcal{A}^c$  is a solution of (21) if and only if, for all  $j \in \mathcal{A}$ ,

$$\begin{aligned} G_j(\hat{\theta}) &= -\text{sign}(\hat{\theta}_j)\mu \text{ if } \hat{\theta}_j \neq 0 \\ |G_j(\hat{\theta})| &\leq \mu \text{ if } \hat{\theta}_j = 0 \end{aligned} \quad (26)$$

Moreover, if the solution is not unique and  $|G_j(\hat{\theta})| < \mu$  for some solution  $\hat{\theta}$ , then  $\hat{\theta}_j = 0$  for all solutions of (21).

**Proposition 3 :** Let  $\hat{\theta}^{\mathcal{A},\mu}$  be defined as in (21). Then, under the assumptions of *Theorem 1*, for any constant  $c_3 > 1$ , for large enough  $n$ ,

$$\mathbb{P}(\text{sign}(\hat{\theta}_j^{\mathcal{A},\mu}) = \text{sign}(\theta_j), \text{ for all } j \in \mathcal{A}) \geq 1 - O(n^{-c_3}). \quad (27)$$

**Lemma 2 :** Define

$$\mathcal{E}_{\mathcal{B},\mu} = \{ \max_{j \in \mathcal{A}^c \cap \mathcal{B}} |G_j(\hat{\theta}^{\mathcal{A},\mu})| < \mu \} \cap \{ \hat{\mathcal{B}} = \mathcal{B} \} \quad (28)$$

On  $\mathcal{E}_{\mathcal{B},\mu}$ ,  $\hat{\theta}^{\mathcal{B},\mu}$  is the unique solution of (20) and  $\hat{\theta}^{\mathcal{A},\mu}$  is the unique solution of (21), and  $\hat{\theta}^{\mathcal{B},\mu} = \hat{\theta}^{\mathcal{A},\mu}$ . Also, under the assumptions of *Theorem 1*, there exists a  $c_4 > 1$  such that, for large enough  $n$ ,

$$\mathbb{P}(\mathcal{E}_{\mathcal{B},\mu}^c) = O(n^{-c_4}). \quad (29)$$

Further, if we define

$$\mathcal{E}_\mu = \{\max_{j \in \mathcal{A}^c} |G_j(\widehat{\theta}^{\mathcal{A}, \mu})| < \mu\} \cap \{\widehat{\mathcal{B}} = \mathcal{B}\}, \quad (30)$$

then under the extra assumption that  $\frac{q_n \log p_n}{n} = o(1)$ , (29) holds with  $\mathcal{E}_{\mathcal{B}, \mu}$  replaced by  $\mathcal{E}_\mu$ . On  $\mathcal{E}_\mu$ ,  $\widehat{\theta}^\mu$  is the unique solution of (19) and  $\widehat{\theta}^\mu = \widehat{\theta}^{\widehat{\mathcal{B}}, \mu} = \widehat{\theta}^{\mathcal{A}, \mu}$ .

#### 4.10 Effect of projection

An important consequence of the projection is that the measurement noise  $Z$  is projected onto a  $K$  dimensional space (that under our assumptions also contains the important components of the predictors of  $Y$ ). This results in a stable behavior of the residual of the projected response  $\Delta$  given by

$$\Delta := \widetilde{Y} - \mathbf{X}\theta = \widetilde{Y} - \mathbf{X}_{\mathcal{A}}\theta_{\mathcal{A}}. \quad (31)$$

even as dimension  $p_n$  becomes large. This can be stated formally in the following proposition.

**Proposition 4 :** Suppose that assumptions **A1-A5** hold. Then there is a constant  $\gamma_3 := \gamma_3(\sigma_0, \lambda_1, \dots, \lambda_K + 1)$ , such that for any  $c_6 > 1$  there exists a constant  $d_6 > 0$  so that, for large enough  $n$ ,

$$\mathbb{P}(\|\Delta\|_2 \leq d_6(\gamma_3 \sqrt{\bar{q}_n \vee \log n} + \sigma_1 \sqrt{K \log n})) \geq 1 - n^{-c_6}. \quad (32)$$

As a direct corollary to this we have the following result about the risk behavior of the OLS-estimator (under  $L^2$  loss) of the preconditioned data

after we have selected the variables by solving the optimization problem (20).

**Corollary 1 :** Suppose that conditions of *Theorem 1* hold. Then for any  $c_7 \geq 1$ , there is  $d_7 > 0$  such that

$$\mathbb{P}(\|\hat{\theta}^{\hat{\mathcal{A}}_{\hat{\mathcal{B}},\mu},OLS} - \theta\|_2 \leq d_7 \sigma_0^{-1} (\gamma_3 \sqrt{\frac{\bar{q}_n \vee \log n}{n}} + \sigma_1 \sqrt{\frac{K \log n}{n}})) \geq 1 - n^{-c_7}, \quad (33)$$

where  $\hat{\theta}^{\hat{\mathcal{A}}_{\hat{\mathcal{B}},\mu},OLS} = (\mathbf{X}_{\hat{\mathcal{A}}}^T \mathbf{X}_{\hat{\mathcal{A}}})^{-1} \mathbf{X}_{\hat{\mathcal{A}}}^T \tilde{Y}$ , and  $\hat{\mathcal{A}} = \hat{\mathcal{A}}_{\hat{\mathcal{B}},\mu_n} = \{j \in \mathcal{P} : \hat{\theta}_{\hat{\mathcal{B}},\mu_n}^j \neq 0\}$ .

As a comparison we can think of the situation when  $\mathcal{A}$  is actually known, and consider the  $L^2$  risk behavior of the OLS estimator restricted only to the subset of variables  $\mathcal{A}$ . Then  $\hat{\theta}^{\mathcal{A},OLS} = (\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1} \mathbf{X}_{\mathcal{A}}^T Y$ . Using the fact that conditional on  $\mathbf{X}_{\mathcal{A}}$ ,  $\hat{\theta}_{\mathcal{A}}^{\mathcal{A},OLS}$  has  $N(\theta_{\mathcal{A}}, \sigma_\varepsilon^2 (\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1})$  distribution, and the fact that the smallest eigenvalue of  $\Sigma_{\mathcal{A}\mathcal{A}}^{-1}$  is at least  $\ell_1^{-1}$ , it follows (using *Lemma A.1*) that there is a constant  $d'_7 > 0$  such that

$$\mathbb{P}(\|\hat{\theta}^{\mathcal{A},OLS} - \theta\|_2 \geq d'_7 \ell_1^{-1/2} \sigma_\varepsilon \sqrt{\frac{q_n}{n}}) \geq 1 - n^{-c_7}. \quad (34)$$

Comparing (34) with (33), we see that if  $q_n \gg \log n$  and  $\sigma_1 \gg \sqrt{\bar{q}_n/q_n}$ , the estimator  $\hat{\theta}^{\hat{\mathcal{A}}_{\hat{\mathcal{B}},\mu},OLS}$  has better risk performance than  $\hat{\theta}^{\mathcal{A},OLS}$ .

As a remark, we point out that the bound in (33) can be improved under specific circumstances (e.g. when  $\bar{\delta}_n$ , the ‘‘selection bias’’ term defined in **A5**, is of a smaller order) by carrying out a second order analysis of the eigenvectors  $\{\tilde{\mathbf{u}}_k\}_{k=1}^K$  (see Appendix of Bair et al. (2006)). The same holds

for the bounds on the partial correlations  $\frac{1}{n}\langle(I - P_{\mathbf{X}_{\mathcal{A}}})X_j, \tilde{Y}\rangle$ , for  $j \in \mathcal{A}^c$ , given the “signal” variables  $\{X_l : l \in \mathcal{A}\}$ , that are needed in the proof of *Proposition 3* and *Lemma 2*. However, the result is given here just to emphasize the point that preconditioning stabilizes the fluctuation in  $\tilde{Y} - \mathbf{X}\theta$ , and so, partly to keep the exposition brief, we do not present the somewhat tedious and technical work needed to carry out such an analysis.

As a further comparison, we consider the contribution of the measurement noise  $Z$  in the maximal empirical partial correlation  $\max_{j \in \mathcal{A}^c} |\frac{1}{n}\langle(I - P_{\mathbf{X}_{\mathcal{A}}})X_j, \tilde{Y}\rangle|$ , given  $\{X_l : l \in \mathcal{A}\}$ . For the pre-conditioned response this contribution is (with probability at least  $1 - O(n^{-c})$  for some  $c > 1$ ) of the order  $O(\frac{\sigma_1\sqrt{\log n}}{\sqrt{n}})$ , instead of  $O(\frac{\sigma_1\sqrt{\log p_n}}{\sqrt{n}})$  as would be the case if one uses  $Y$  instead of  $\tilde{Y}$ . So, if  $\log p_n \gg \log n$ , then the contribution is smaller for the pre-conditioned response. Formalizing this argument, we derive the following asymptotic result about the model selection property of LASSO estimator that clearly indicates that under latter circumstances SPC + LASSO procedure can outperform conventional LASSO in terms of variable selection.

**Proposition 5 :** Suppose that  $\log p_n = cn^\alpha$  for some  $\alpha \in (0, 1)$  and some  $c > 0$ . Suppose that  $\mathcal{A} = \mathcal{A}_+ \cup \mathcal{A}_-$ , with  $\mathcal{A}_+$  and  $\mathcal{A}_-$  disjoint and  $\mathcal{A}_-$  nonempty such that  $\|\theta_{\mathcal{A}_-}\|_2 = o(n^{-(1-\alpha)/2})$ . Assume that  $M = K$ ,  $\mathcal{B} = \mathcal{D}$  (so that for all  $j \notin \mathcal{B}$ ,  $X_j$  are i.i.d.  $N(0, \sigma_0^2)$ ), and  $\sigma_1$  is fixed . Suppose further that all the assumptions of *Theorem 1* hold, and there is a  $\delta_+ \in (0, 1)$  such

that (if  $\mathcal{A}_+$  is nonempty)

$$\max_{j \notin \mathcal{A}_+} \|\Sigma_{\mathcal{A}_+ \mathcal{A}_+}^{-1} \Sigma_{\mathcal{A}_+ j}\|_1 < \delta_+. \quad (35)$$

Then, given  $c_8 \geq 1$ , for all  $\mu_n \geq 0$ , for large enough  $n$ ,

$$\mathbb{P}(\widehat{\mathcal{A}}_{\mu_n}^{LASSO} \neq \mathcal{A}) \geq 1 - n^{-c_8}, \quad (36)$$

where  $\widehat{\mathcal{A}}_{\mu_n}^{LASSO} = \{j \in \mathcal{P} : \widehat{\theta}_j^{LASSO, \mu_n} \neq 0\}$ , where

$$\widehat{\theta}^{LASSO, \mu_n} = \arg \min_{\zeta \in \mathbb{R}^p} \frac{1}{n} \|Y - \mathbf{X}\zeta\|_2^2 + \mu_n \|\zeta\|_1. \quad (37)$$

Proposition 5 shows that if  $\alpha > 1 - 2\kappa_2$ , so that  $\eta_n = o(n^{-(1-\alpha)/2})$ , and the assumptions of Proposition 5 are satisfied, then the SPC + LASSO approach (solving the optimization problem (20) or (19)) can identify  $\mathcal{A}$  with appropriate choice of penalization parameter  $\mu_n$  (as indicated in *Theorem 1*) while LASSO cannot, with any choice of the penalty parameter.

## 5 Classification problems and further topics

The pre-conditioning idea has potential application in any supervised learning problem in which the number of features greatly exceeds the number of observations. A key component is the availability of a consistent estimator for the construction of the pre-conditioned outcome variable.

For example, pre-conditioning can be applied to classification problems. Conceptually, we separate the problems of a) obtaining a good classifier and

b) selecting a small set of good features for classification. Many classifiers, such as the support vector machine, are effective at finding a good separator for the classes. However they are much less effective in distilling these features down into a smaller set of uncorrelated features.

Consider a two-class problem, and suppose we have trained a classifier, yielding estimates  $\hat{p}_i$ , the probability of class 2 for observation  $i = 1, 2, \dots, N$ . Then in the second stage, we apply a selection procedure such as forward stepwise or the LASSO, to an appropriate function of  $\hat{p}_i$ ; the quantity  $\log[\hat{p}_i/(1 - \hat{p}_i)]$  is a logical choice.

We generated data as in example of section 3; however we turned it into a classification problem by defining the outcome class  $g_i$  as 1 if  $y_i < 0$  and 2 otherwise. We applied the nearest shrunken centroid (NSC) classifier of Tibshirani et al. (2001), a method for classifying microarray samples. We applied forward stepwise regression both to  $g_i$  directly (labeled FS), and to the output  $\log(\hat{p}_i/(1 - \hat{p}_i))$  of the NSC classifier (labeled NSC/FS).

The results of 10 simulations are shown in Figure 5. We see that NSC/FS does not improve the test error of FS, but as shown in the bottom left panel, it does increase the number of “good” predictors that are found. This is a topic of further study.

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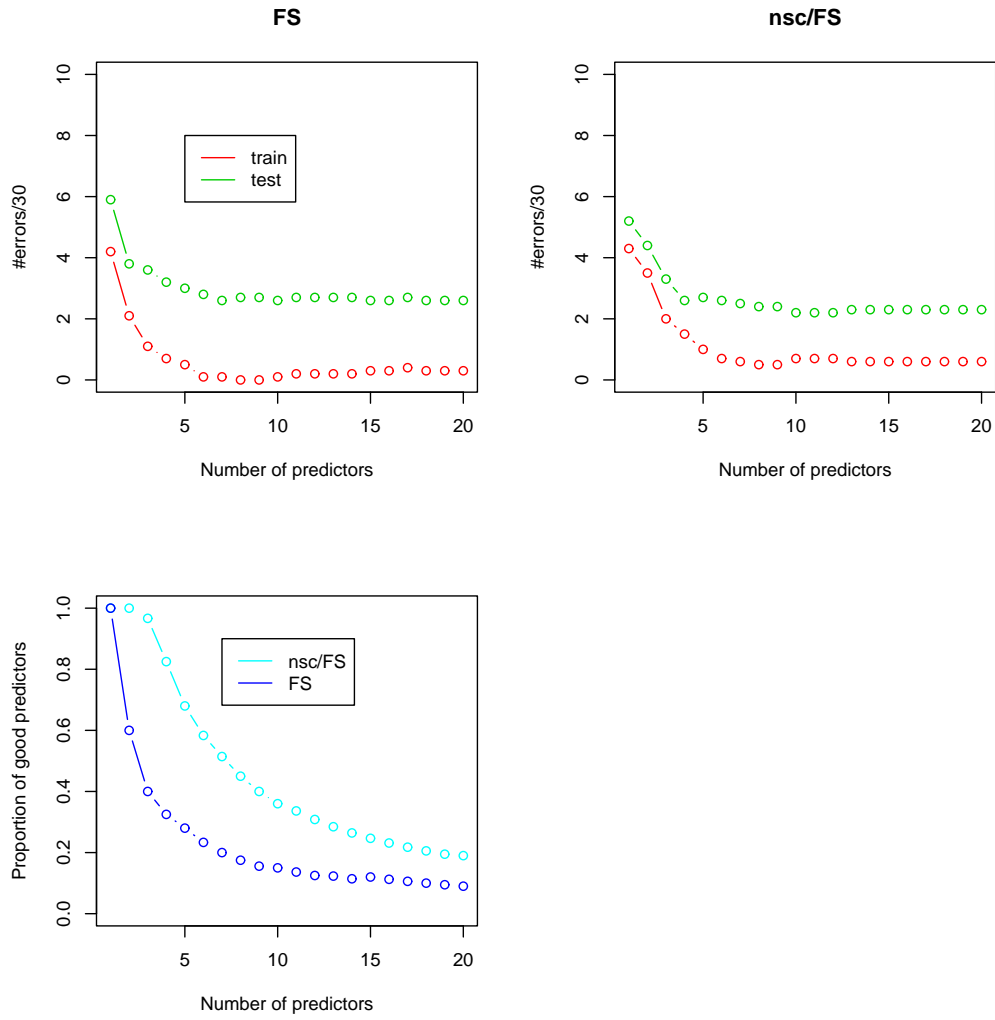


Figure 5: Results of applying pre-conditioning in a classification setting. Top left panel shows the number of test misclassification errors from forward stepwise regression; in the top right panel we have applied forward stepwise regression to the pre-conditioned estimates from nearest shrunken centroid classifier. The proportion of good predictors selected by each method is shown in the bottom left.



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## Appendix

The following lemma about the extreme eigenvalues of a Wishart( $N, n$ ) matrix will be very helpful for a lot of computations that follow. This is a simple corollary of Theorem 2.13 of ? and a proof can be found in Paul (2005).

**Lemma A.1 :** Let  $\mathbf{S} = \frac{1}{n}WW^T$  where  $W$  is an  $N \times n$  matrix with i.i.d.  $N(0, 1)$  entries. Assume that  $1 \leq N < n$ . Let  $m_1(N, n) := (1 + \sqrt{\frac{N}{n}})^2$  and  $m_N(N, n) := (1 - \sqrt{\frac{N}{n}})^2$ . Let  $\lambda_1(\mathbf{S})$  and  $\lambda_N(\mathbf{S})$  denote the largest and the smallest eigenvalues of  $\mathbf{S}$ . Then, for  $t > 0$ ,

$$\begin{aligned} \mathbb{P}(\lambda_1(\mathbf{S}) - m_1(N, n) > t) &\leq \exp\left(-\frac{n}{2}(\sqrt{t + m_1(N, n)} - \sqrt{m_1(N, n)})^2\right) \\ &\leq \exp\left(-\frac{nt^2}{4(t + 2m_1(N, n))}\right), \end{aligned} \quad (38)$$

and

$$\begin{aligned} \mathbb{P}(\lambda_N(\mathbf{S}) - m_N(N, n) < -t) &\leq \exp\left(-\frac{n}{2}(\sqrt{t + m_N(N, n)} - \sqrt{m_N(N, n)})^2\right) \\ &\leq \exp\left(-\frac{nt^2}{4(t + 2m_N(N, n))}\right). \end{aligned} \quad (39)$$

This lemma is typically applied to control the fluctuations of  $\lambda_{\max}(\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D)$ , and  $\lambda_{\min}(\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D)$ , the largest and smallest eigenvalues of  $\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D$  for some set  $D \subset \mathbb{R}^p$  with  $|D| < n$ . Then we shall use the fact that in distribution,

$$\lambda_{\min}(\Sigma_{DD})\lambda_{|D|}(\mathbf{S}_D) \leq \lambda_{\min}(\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D) \leq \lambda_{\max}(\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D) \leq \lambda_{\max}(\Sigma_{DD})\lambda_1(\mathbf{S}_D)$$

where  $n\mathbf{S}_D$  has a Wishart( $n, |D|$ ) distribution. Now we can use the bounds (38) and (39), with  $N = |D|$  to provide probabilistic lower and upper bounds on the two extreme eigenvalues of  $\frac{1}{n}\mathbf{X}_D^T\mathbf{X}_D$ .

**Proof of Proposition 1 :** Let us denote the empirical correlation between  $X_j$  and  $Y$  by  $\text{corr}(X_j, Y)$ . Then

$$\text{corr}(X_j, Y) = \frac{\frac{1}{n}\langle X_j, Y \rangle}{\frac{1}{n} \| X_j \|_2 \| Y \|_2}$$

Since  $\| X_j \|^2 / \Sigma_{jj}$  and  $\| Y \|^2 / \sigma_{yy}$  have  $\chi_n^2$  distributions, given any constant  $c_1 > 0$ , there are  $d_1, d_2 > 0$  such that with probability at least  $1 - n^{-c_1}$ , we have  $|\frac{1}{n} \| Y \|^2 - \sigma_{yy}| \leq \sigma_{yy}d_1\sqrt{\frac{\log n}{n}}$ , and for all  $j \in \mathcal{P}$ ,  $|\frac{1}{n} \| X_j \|^2 - \Sigma_{jj}| \leq \Sigma_{jj}d_2\sqrt{\frac{\log(pn)}{n}}$ . Next,  $\langle X_j, Y \rangle - n\Sigma_{jy}$  is a sum of i.i.d. random variables, say  $\{\xi_{i,jy} : i = 1, \dots, n\}$ , where  $\xi_{i,jy}$ 's have 0 mean

and covariance  $\alpha_{jy} = \sigma_{yy}\Sigma_{jj} + \Sigma_{jy}^2$ . Therefore,  $\alpha_{jy} \leq 2\sigma_{yy}\Sigma_{jj}$ . Thus, by Bernstein's inequality, with probability at least  $1 - n^{-c_1}$ , for all  $j \in \mathcal{P}$ ,  $|\frac{1}{n}\langle X_j, Y \rangle - \Sigma_{jy}| \leq d_3\sqrt{\sigma_{yy}}\sqrt{\Sigma_{jj}}\sqrt{\frac{\log(np_n)}{n}}$ , for some constant  $d_3$ . Observe that  $X_j$  and  $Y$  are independent for  $j \notin \mathcal{B}$ , so that  $\Sigma_{jy} = 0$  for all  $j \in \mathcal{B}^c$ . The proof follows by invoking assumption **A4**.

**Proof of Lemma 1 :** Same as proof of *Lemma 1* in Meinshausen & Bühlmann (2006).

**Proof of Theorem 1 :** First consider  $\widehat{\mathcal{A}}_{\widehat{\mathcal{B}}, \mu_n}$ . Observe that

$$\mathbb{P}(\widehat{\mathcal{A}}_{\widehat{\mathcal{B}}, \mu_n} \subset \mathcal{A}) = 1 - \mathbb{P}(\exists j \in \mathcal{A}^c \cap \widehat{\mathcal{B}} \text{ such that } \widehat{\theta}_j^{\widehat{\mathcal{B}}, \mu_n} \neq 0).$$

From *Lemma 2*, on  $\mathcal{E}_{\mathcal{B}, \mu}$ ,  $\widehat{\theta}^{\widehat{\mathcal{B}}, \mu} = \widehat{\theta}^{\mathcal{A}, \mu}$  is the unique solution of (20) and (21).

Therefore,

$$\mathbb{P}(\exists j \in \mathcal{A}^c \cap \widehat{\mathcal{B}} \text{ such that } \widehat{\theta}_j^{\widehat{\mathcal{B}}, \mu_n} \neq 0) \leq 1 - \mathbb{P}(\mathcal{E}_{\mathcal{B}, \mu_n}).$$

Now apply (29) to conclude the proof. By second part of *Lemma 2*, one can replace  $\widehat{\theta}^{\widehat{\mathcal{B}}, \mu_n}$ ,  $\widehat{\mathcal{A}}_{\widehat{\mathcal{B}}, \mu_n}$  and  $\mathcal{E}_{\mathcal{B}, \mu_n}$  by  $\widehat{\theta}^{\mu_n}$ ,  $\widehat{\mathcal{A}}_{\mu_n}$  and  $\mathcal{E}_{\mu_n}$ , respectively.

**Proof of Theorem 2 :** Observe that,

$$\mathbb{P}(\mathcal{A} \subset \widehat{\mathcal{A}}_{\widehat{\mathcal{B}}, \mu}, \widehat{\mathcal{B}} = \mathcal{B}) = \mathbb{P}(\widehat{\mathcal{B}} = \mathcal{B}) - \mathbb{P}(\exists j \in \mathcal{A} : \widehat{\theta}_j^{\widehat{\mathcal{B}}, \mu} = 0, \widehat{\mathcal{B}} = \mathcal{B}).$$

Since, by *Lemma 2*, on  $\mathcal{E}_{\mathcal{B}, \mu}$ ,  $\widehat{\theta}^{\mathcal{A}, \mu} = \widehat{\theta}^{\widehat{\mathcal{B}}, \mu}$ , it follows that

$$\mathbb{P}(\exists j \in \mathcal{A} : \widehat{\theta}_j^{\widehat{\mathcal{B}}, \mu} = 0, \widehat{\mathcal{B}} = \mathcal{B}) \leq \mathbb{P}(\exists j \in \mathcal{A} : \widehat{\theta}_j^{\mathcal{A}, \mu} = 0) + \mathbb{P}(\mathcal{E}_{\mathcal{B}, \mu}^c).$$

Now, the proof is completed by applying *Proposition 1*, (29) and (27). Proof for  $\widehat{\mathcal{A}}_{\mu_n}$  is analogous.

**Proof of Lemma 2 :** From the definition (21) and *Lemma 1*, it follows that (noticing that  $\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}}$  is invertible almost surely) on the set

$$B_{\mu} := \{\text{sign}(\widehat{\theta}_j^{\mathcal{A},\mu}) = \text{sign}(\theta_j) \quad \text{for all } j \in \mathcal{A}\} \quad (40)$$

$$\widehat{\theta}_{\mathcal{A}}^{\mathcal{A},\mu} = \left(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}}\right)^{-1} \left(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \widetilde{Y}\right) - \frac{\mu}{2} \left(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}}\right)^{-1} \text{sign}(\theta_{\mathcal{A}}). \quad (41)$$

On the other hand, for all  $j$ ,

$$G_j(\widehat{\theta}^{\mathcal{A},\mu}) = -\frac{2}{n} \langle \widetilde{Y} - \mathbf{X} \widehat{\theta}^{\mathcal{A},\mu}, X_j \rangle = -\frac{2}{n} \langle \widetilde{Y} - \mathbf{X}_{\mathcal{A}} \widehat{\theta}_{\mathcal{A}}^{\mathcal{A},\mu}, X_j \rangle$$

since, by definition,  $\widehat{\theta}_{\mathcal{A}^c}^{\mathcal{A},\mu} = 0$ . Therefore it follows from (41) and the definition of  $\Delta$  in (31) that, on  $B_{\mu}$ ,

$$\widetilde{Y} - \mathbf{X} \widehat{\theta}^{\mathcal{A},\mu} = (I - P_{\mathbf{X}_{\mathcal{A}}}) \Delta - \frac{\mu}{2} \mathbf{X}_{\mathcal{A}} \left(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}}\right)^{-1} \text{sign}(\theta_{\mathcal{A}}) \quad (42)$$

The matrix  $P_{\mathbf{X}_{\mathcal{A}}} := \mathbf{X}_{\mathcal{A}} (\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1} \mathbf{X}_{\mathcal{A}}^T$  is the orthogonal projection matrix onto the column space of  $\mathbf{X}_{\mathcal{A}}$ .

Then, since  $\|\Delta\|_2 = O(\sigma_0 \sqrt{q_n} \sqrt{\log n} + \sigma_1 \sqrt{K} \sqrt{\log n}) = o(\sqrt{n} \mu_n)$  with probability at least  $1 - n^{-c_4}$ , for some  $c_4 > 1$  (by *Proposition 4*), and  $\|X_j\|_2^2 \leq n \Sigma_{jj} (1 + d_4 \sqrt{\frac{\log(pn)}{n}})$  with probability at least  $1 - p_n^{-1} n^{-c_4}$ , it follows that except on a set of probability  $O(n^{-c_4})$ ,

$$\max_{j \in \mathcal{A}^c} \left| \frac{1}{n} X_j^T (I - P_{\mathbf{X}_{\mathcal{A}}}) \Delta \right| \leq \max_{j \in \mathcal{A}^c} \frac{1}{n} \|X_j\|_2 \|\Delta\|_2 \leq \frac{\epsilon \mu_n}{2}, \quad (43)$$

for some  $\epsilon > 0$  satisfying  $2\epsilon < 1 - \delta$  (where  $\delta$  is as in assumption **A3**).

Then, by observing that

$$X_j^T \mathbf{X}_A (\mathbf{X}_A^T \mathbf{X}_A)^{-1} - \Sigma_{jA} \Sigma_{AA}^{-1} = (X_j^T - \Sigma_{jA} \Sigma_{AA}^{-1} \mathbf{X}_A^T) \mathbf{X}_A (\mathbf{X}_A^T \mathbf{X}_A)^{-1},$$

recalling that  $X_j^T - \Sigma_{jA} \Sigma_{AA}^{-1} \mathbf{X}_A^T$  is independent of  $\mathbf{X}_A$  and has i.i.d.  $N(0, \Sigma_{jj \cdot A})$  entries, where  $\Sigma_{jj \cdot A} = \Sigma_{jj} - \Sigma_{jA} \Sigma_{AA}^{-1} \Sigma_{Aj}$ , so that

$$(X_j^T - \Sigma_{jA} \Sigma_{AA}^{-1} \mathbf{X}_A^T) \mathbf{X}_A (\mathbf{X}_A^T \mathbf{X}_A)^{-1} \text{sign}(\theta_A) \mid \mathbf{X}_A \sim N(0, \Sigma_{jj \cdot A} \text{sign}(\theta_A)^T (\mathbf{X}_A^T \mathbf{X}_A)^{-1} \text{sign}(\theta_A)),$$

using the fact that  $\| \frac{1}{n} \mathbf{X}_A^T \mathbf{X}_A - \Sigma_{AA} \| \leq \ell_1 (2\sqrt{\frac{q_n}{n}} + \frac{q_n}{n} + O(\sqrt{\frac{\log n}{n}}))$  with probability at least  $1 - n^{-c_4}$  (by *Lemma A.1*), and noticing that  $\| \text{sign}(\theta_A) \|_2 = \sqrt{q_n}$ , so that

$$\text{sign}(\theta_A)^T (\mathbf{X}_A^T \mathbf{X}_A)^{-1} \text{sign}(\theta_A) \leq \frac{q_n}{n} \left\| \left( \frac{1}{n} \mathbf{X}_A^T \mathbf{X}_A \right)^{-1} \right\| \leq \frac{q_n}{n} \frac{\| \Sigma_{AA}^{-1} \|}{1 - \| \Sigma_{AA}^{-1} \| \| \frac{1}{n} \mathbf{X}_A^T \mathbf{X}_A - \Sigma_{AA} \|}, \quad (44)$$

as long as the denominator is positive, and using the inequality  $1 - \Phi(x) \leq \frac{1}{x} \phi(x)$ , for  $x > 0$  and  $\Phi(\cdot)$  and  $\phi(\cdot)$  denoting  $N(0, 1)$  c.d.f. and p.d.f. respectively, we conclude that except on a set of probability  $O(n^{-c_4})$ ,

$$\max_{j \in \mathcal{A}^c \cap \mathcal{B}} \left| \left( \frac{1}{n} X_j^T \mathbf{X}_A \left( \frac{1}{n} \mathbf{X}_A^T \mathbf{X}_A \right)^{-1} - \Sigma_{jA} \Sigma_{AA}^{-1} \right) \text{sign}(\theta_A) \right| < \epsilon. \quad (45)$$

Note also that, if  $p_n$  is such that  $\frac{q_n \log p_n}{n} = o(1)$ , then the same conclusion as (45) holds even if we replace the maximum over the set  $\mathcal{A}^c \cap \mathcal{B}$  by the maximum over the set  $\mathcal{A}^c$ . From (43), (45) and *Proposition 3* the result

follows, since by **A7**,

$$\max_{j \in \mathcal{A}^c} |\Sigma_{j\mathcal{A}} \Sigma_{\mathcal{A}\mathcal{A}}^{-1} \text{sign}(\theta_{\mathcal{A}})| < \delta.$$

**Proof of Proposition 3 :** The proof is by a variational argument. Let  $\widehat{\theta}$  be any vector with  $\text{sign}(\widehat{\theta}_j) \neq \text{sign}(\theta_j)$  for some  $j \in \mathcal{A}$ . We show that with very high probability such a vector cannot be a solution of the optimization problem (21).

First, we use an argument, analogous to the proof of Theorem 1 in ?, to claim that we can restrict ourselves to the case when  $\|\widehat{\theta} - \theta\|_2 = O(\sqrt{q_n \mu_n})$ . Accordingly, we show that for a large constant  $C$ , for any  $c_5 > 1$ , and for large enough  $n$ ,

$$\mathbb{P}\left(\sup_{\mathbf{a} \in \mathbb{R}^p: \mathbf{a}_{\mathcal{A}^c} = 0, \|\mathbf{a}\|_2 = C} Q_n(\theta + \sqrt{q_n \mu_n} \mathbf{a}) < Q_n(\theta)\right) \geq 1 - n^{-c_5}, \quad (46)$$

where  $Q_n(\zeta) = \frac{1}{n} \|\widetilde{Y} - \mathbf{X}\zeta\|_2^2 + \mu_n \|\zeta\|_1$ . (46) implies that there exists a solution of the optimization problem (with probability at least  $1 - n^{-c_5}$ ) in the set  $\{\theta + \mathbf{a} : \mathbf{a}_{\mathcal{A}^c} = 0, \|\mathbf{a}\|_2 = C\}$ . By strict convexity of the objective function  $Q_n(\cdot)$ , this implies our claim.

To prove (46), observe that for any  $\mathbf{a}$  satisfying the conditions,

$$\begin{aligned} Q_n(\theta) - Q_n(\theta + \sqrt{q_n \mu_n} \mathbf{a}) &\leq -q_n \mu_n^2 \frac{1}{n} \|\mathbf{X}_{\mathcal{A}} \mathbf{a}_{\mathcal{A}}\|_2^2 - \sqrt{q_n \mu_n} \frac{2}{n} \langle \Delta, \mathbf{X}_{\mathcal{A}} \mathbf{a}_{\mathcal{A}} \rangle + \sqrt{q_n \mu_n}^2 \|\mathbf{a}_{\mathcal{A}}\|_1 \\ &\leq -q_n \mu_n^2 \left( \frac{1}{n} \|\mathbf{X}_{\mathcal{A}} \mathbf{a}_{\mathcal{A}}\|_2^2 - \|\mathbf{a}_{\mathcal{A}}\|_2 \right) + \sqrt{q_n \mu_n} \frac{2}{n} \|\Delta\|_2 \|\mathbf{X}_{\mathcal{A}} \mathbf{a}_{\mathcal{A}}\|_2, \end{aligned}$$

where in the second step we used the fact that  $\| \mathbf{a}_{\mathcal{A}} \|_1 \leq \sqrt{q_n} \| \mathbf{a}_{\mathcal{A}} \|_2$  since  $|\mathcal{A}| = q_n$ . Now, notice that since  $\| \mathbf{a}_{\mathcal{A}} \|_2 = C$ ,

$$C \sqrt{\lambda_{\min}(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})} \leq \frac{1}{\sqrt{n}} \| \mathbf{X}_{\mathcal{A}} \mathbf{a}_{\mathcal{A}} \|_2 \leq C \sqrt{\lambda_{\max}(\frac{1}{n} \mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})},$$

where  $\lambda_{\min}$  and  $\lambda_{\max}$  are the smallest and largest eigenvalues, respectively. Then observing that  $\sigma_0^2 \leq \lambda_{\min}(\Sigma_{\mathcal{A}\mathcal{A}}) \leq \lambda_{\max}(\Sigma_{\mathcal{A}\mathcal{A}}) \leq \ell_1$ , then applying *Lemma A.1*, and finally noticing that by *Proposition 4*,  $\| \Delta \|_2 = o(\sqrt{n}\mu_n)$  with probability at least  $1 - \frac{1}{2}n^{-c_5}$ , for large enough  $n$ , the proof is completed by choosing  $C$  large enough so that  $C > 2\sigma_0^{-2}$ , say.

So fix  $j \in \mathcal{A}$  and without loss of generality suppose that  $\theta_j > 0$ . Consider any vector  $\hat{\theta}$  such that  $\hat{\theta}_j \leq 0$ , and  $\| \hat{\theta} - \theta \|_2 \leq C_1 \sqrt{q_n} \mu_n$ , for some large constant  $C_1$ . Let  $\mathcal{A}_j := \mathcal{A} \setminus \{j\}$ . Let  $\psi_n > 0$  be a free parameter such that  $\psi_n = o(\mu_n)$ . Define a vector  $\tilde{\theta}$  as follows.  $\tilde{\theta}_{\mathcal{A}_j} = \hat{\theta}_{\mathcal{A}_j}^{\mu_n} - \psi_n \xi^j$  where  $\xi^j := \Sigma_{\mathcal{A}_j \mathcal{A}_j}^{-1} \Sigma_{\mathcal{A}_j j}$ , and  $\tilde{\theta}_j = \hat{\theta}_j + \psi_n$ . Also,  $\tilde{\theta}_{\mathcal{A}^c} = 0$ . Let  $P_{\mathbf{X}_{\mathcal{A}_j}}$  be the projection matrix on the column space of  $\mathbf{X}_{\mathcal{A}_j}$ . Let  $w_j = X_j - \mathbf{X}_{\mathcal{A}_j} \xi^j$ . Then

$$\begin{aligned} & \left[ \frac{1}{n} \| \tilde{Y} - \mathbf{X} \tilde{\theta} \|_2^2 + \mu \| \tilde{\theta} \|_1 \right] - \left[ \frac{1}{n} \| \tilde{Y} - \mathbf{X} \hat{\theta} \|_2^2 + \mu \| \hat{\theta} \|_1 \right] \\ &= \frac{1}{n} \| \tilde{Y} - \mathbf{X}_{\mathcal{A}} \hat{\theta}_{\mathcal{A}} - \psi_n w_j \|_2^2 + \mu_n \| \tilde{\theta}_{\mathcal{A}} \|_1 - \left[ \frac{1}{n} \| \tilde{Y} - \mathbf{X} \hat{\theta} \|_2^2 + \mu \| \hat{\theta} \|_1 \right] \\ &\leq \psi_n^2 \frac{1}{n} \| w_j \|_2^2 - 2\psi_n \frac{1}{n} w_j^T \Delta - 2\psi_n (\theta_j - \hat{\theta}_j) \frac{1}{n} w_j^T X_j \\ &\quad - 2\psi_n \frac{1}{n} w_j^T \mathbf{X}_{\mathcal{A}_j} (\theta_{\mathcal{A}_j} - \hat{\theta}_{\mathcal{A}_j}) + \mu_n \psi_n (1 + \| \xi^j \|_1) \\ &= -2\psi_n (\theta_j - \hat{\theta}_j) \frac{1}{n} \| w_j \|_2^2 + \psi_n^2 \frac{1}{n} \| w_j \|_2^2 - 2\psi_n \frac{1}{n} w_j^T \Delta \\ &\quad - 2\psi_n \frac{1}{n} w_j^T \mathbf{X}_{\mathcal{A}_j} \mathbf{b}^j + \mu_n \psi_n (1 + \| \xi^j \|_1), \end{aligned} \tag{47}$$

where  $\mathbf{b}_j = (\theta_{\mathcal{A}_j} - \widehat{\theta}_{\mathcal{A}_j})\xi^j + (\theta_{\mathcal{A}_j} - \widehat{\theta}_{\mathcal{A}_j})$ . Therefore,  $\|\mathbf{b}^j\|_2 \leq \|\widehat{\theta} - \theta\|_2 (1 + \|\xi^j\|_2)$ . Observe that  $w_j$  is independent of  $\mathbf{X}_{\mathcal{A}_j}$  and has i.i.d.  $N(0, \Sigma_{jj \cdot \mathcal{A}_j})$  entries, so that, the conditional distribution of  $w_j^T \mathbf{X}_{\mathcal{A}_j} \mathbf{b}^j$  given  $\mathbf{X}_{\mathcal{A}_j}$  is  $N(0, \Sigma_{jj \cdot \mathcal{A}_j} \|\mathbf{X}_{\mathcal{A}_j} \mathbf{b}^j\|^2)$ . Therefore, it follows (using *Lemma A.1*) that, for any  $c_5 \geq 2$ , with probability at least  $1 - n^{-c_5}$ , for some constant  $d_5 > 0$ ,

$$\left| \frac{1}{n} w_j^T \mathbf{X}_{\mathcal{A}_j} \mathbf{b}^j \right| \leq d_5 \Sigma_{jj}^{1/2} \ell_1^{1/2} (1 + \|\xi^j\|_2) \sqrt{\frac{\log n}{n}} \|\widehat{\theta} - \theta\|_2.$$

Since  $\|\widehat{\theta} - \theta\|_2 = O(\sqrt{q_n} \mu_n)$ , it follows that with probability at least  $1 - n^{-c_5}$ ,  $|\frac{1}{n} w_j^T \mathbf{X}_{\mathcal{A}_j} \mathbf{b}^j| = o(\mu_n)$ . Again, with probability at least  $1 - n^{-c_5}$ ,  $|\frac{1}{n} w_j^T \Delta| \leq \frac{\|w_j\|_2}{\sqrt{n}} \frac{\|\Delta\|_2}{\sqrt{n}} \leq \epsilon \mu_n$ , for some  $\epsilon > 0$  small, by using *Proposition 4*. By assumption **A8**,  $\|\xi^j\|_1 \leq \vartheta$ . Now, using the facts that  $\theta_j - \widehat{\theta}_j \geq |\theta_j| \geq \eta_n \gg \mu_n$  (by **A6** and the assumption on  $\mu_n$  in *Theorem 1*), and  $\Sigma_{jj \cdot \mathcal{A}_j} \geq \sigma_0^2$ , and since we can handle the case when  $\theta_j < 0$  and  $\widehat{\theta}_j \geq 0$  simply by changing the sign of  $\psi_n$ , the result follows by taking union bound over  $j \in \mathcal{A}$ .

**Proof of Proposition 4 :** First, we find a bound on the quantities  $\|\bar{\mathbf{u}}_k - \mathbf{u}_k\|_2$  and  $|\bar{\ell}_k - \ell_k|$  for  $k = 1, \dots, K$ , where  $(\bar{\mathbf{u}}_{\mathcal{B}k}, \bar{\ell}_k)_{k=1}^K$  are the first  $K$  eigen-pairs for the matrix  $\Sigma_{\mathcal{B}\mathcal{B}}$ , and  $\bar{\mathbf{u}}_{\mathcal{B}^c k} = 0$  for all  $k$ .

First observe that  $\bar{\ell}_k = \bar{\lambda}_k + \sigma_0^2$ , where  $\bar{\lambda}_k$  is the  $k$ -th largest eigenvalue of the matrix  $\mathbf{W}_{\mathcal{B}} \mathbf{W}_{\mathcal{B}}^T + \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T = \Sigma_{\mathcal{B}\mathcal{B}} - \sigma_0^2 I$ , where  $\widetilde{\mathbf{w}}_j = (\sqrt{\lambda_{K+1}} u_{(K+1)j}, \dots, \sqrt{\lambda_M} u_{Mj})$ . Define  $P_{\mathcal{W}, \mathcal{B}}$  to be the projection matrix onto the column space of  $\mathbf{W}_{\mathcal{B}}$  and  $P_{\mathcal{W}, \mathcal{B}}^\perp = I - P_{\mathcal{W}, \mathcal{B}}$  as the projection matrix on the orthogonal complement of



this space in  $\mathbb{R}^{\mathcal{B}}$ . Then notice that  $(\mathbf{u}_k, \lambda_k)_{k=1}^K$  are the first  $K$  eigen-pairs of the matrix

$$\mathbf{K} = \mathbf{W}\mathbf{W}^T + \begin{bmatrix} P_{\widetilde{W},\mathcal{B}}^\perp \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T P_{\widetilde{W},\mathcal{B}}^\perp & O \\ O & O \end{bmatrix} = \begin{bmatrix} \mathbf{W}_{\mathcal{B}} \mathbf{W}_{\mathcal{B}}^T + P_{\widetilde{W},\mathcal{B}}^\perp \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T P_{\widetilde{W},\mathcal{B}}^\perp & \mathbf{W}_{\mathcal{B}} \mathbf{W}_{\mathcal{B}^c}^T \\ \mathbf{W}_{\mathcal{B}^c} \mathbf{W}_{\mathcal{B}}^T & \mathbf{W}_{\mathcal{B}^c} \mathbf{W}_{\mathcal{B}^c}^T \end{bmatrix}.$$

So we can write

$$\begin{bmatrix} \mathbf{W}_{\mathcal{B}} \mathbf{W}_{\mathcal{B}}^T + \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T & O \\ O & O \end{bmatrix} = \mathbf{K} + \mathbf{R}^{(1)} - \mathbf{R}^{(2)},$$

where

$$\mathbf{R}^{(1)} = \begin{bmatrix} P_{\widetilde{W},\mathcal{B}}^\perp \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T P_{\widetilde{W},\mathcal{B}}^\perp + P_{W,\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}} \widetilde{\mathbf{W}}_{\mathcal{B}}^T P_{W,\mathcal{B}}^\perp & O \\ O & O \end{bmatrix} \text{ and } \mathbf{R}^{(2)} = \begin{bmatrix} O & \mathbf{W}_{\mathcal{B}} \mathbf{W}_{\mathcal{B}^c}^T \\ \mathbf{W}_{\mathcal{B}^c} \mathbf{W}_{\mathcal{B}}^T & \mathbf{W}_{\mathcal{B}^c} \mathbf{W}_{\mathcal{B}^c}^T \end{bmatrix}.$$

A crucial observation is that for  $1 \leq k, l \leq M$ ,

$$|\langle \mathbf{u}_{\mathcal{B}k}, \mathbf{u}_{\mathcal{B}l} \rangle - \delta_{kl}| = |\langle \mathbf{u}_{\mathcal{B}^c k}, \mathbf{u}_{\mathcal{B}^c l} \rangle| \leq \| \mathbf{u}_{\mathcal{B}^c k} \|_2 \| \mathbf{u}_{\mathcal{B}^c l} \|_2$$

because of the orthonormality of the vectors  $\mathbf{u}_1, \dots, \mathbf{u}_M$ . Here  $\delta_{kl}$  is 1 if  $k = l$  and is zero otherwise. Therefore, using matrix perturbation analysis (e.g. Kneip (2001)), one can easily show that

$$\max_{1 \leq k \leq K} |\bar{\lambda}_k - \lambda_k| = O(\bar{\delta}_n), \quad \text{and} \quad \max_{1 \leq k \leq K} \| \bar{\mathbf{u}}_k - \mathbf{u}_k \|_2 = O(\sqrt{\bar{\delta}_n}), \quad (48)$$

where the constants in the  $O()$  terms depend on  $\lambda_1, \dots, \lambda_M$ .

Next, by *Proposition 1*, we can restrict attention to the set  $\{\widehat{\mathcal{B}} = \mathcal{B}\}$ .

Next, for any  $1 \leq k \leq K$  and  $1 \leq l \leq M$ ,

$$|\langle \tilde{\mathbf{u}}_k, \mathbf{u}_l \rangle - \langle \mathbf{u}_k, \mathbf{u}_l \rangle| \leq |\langle \tilde{\mathbf{u}}_k - \bar{\mathbf{u}}_k, \mathbf{u}_l \rangle| + |\langle \bar{\mathbf{u}}_k - \mathbf{u}_k, \mathbf{u}_l \rangle| \leq \| \tilde{\mathbf{u}}_k - \bar{\mathbf{u}}_k \|_2 + \| \bar{\mathbf{u}}_k - \mathbf{u}_k \|_2. \quad (49)$$

Further, by assumption of independent subsamples,  $\{\mathbf{E}\tilde{\mathbf{u}}_k : k = 1, \dots, K\}$  are i.i.d.  $N(0, I_n)$  vectors and are independent of  $\{\mathbf{v}_k : 1 \leq k \leq M\}$ .

Therefore, by properties of Gaussian random vectors, for any  $c > 1$  there is a  $d > 0$  such that for  $n$  large enough,

$$\mathbb{P}\left(\max\left\{\max_{1 \leq k, l \leq K} \left|\frac{1}{n}\langle \mathbf{E}\tilde{\mathbf{u}}_k, \mathbf{v}_l \rangle\right|, \max_{1 \leq k \leq K} \max_{1 \leq l \leq M} \left|\frac{1}{n}\langle \mathbf{v}_k, \mathbf{v}_l \rangle - \delta_{kl}\right|\right\} \leq d\sqrt{\frac{\log n}{n}}\right) \geq 1 - n^{-c}. \quad (50)$$

Again, since  $\tilde{P}_K$  is a projection onto a  $K$  dimensional subspace, and by assumption of independent subsample  $\tilde{P}_K$  and  $Z$  are independent, it follows that conditionally on  $\mathbf{X}$  and the first subsample,  $\tilde{P}_K Z = \frac{1}{n} \sum_{k=1}^K \frac{1}{l_k} \langle \mathbf{X}\tilde{\mathbf{u}}_k, Z \rangle \mathbf{X}\tilde{\mathbf{u}}_k$  has the  $N(0, \tilde{P}_K)$  distribution. This implies that  $\|\tilde{P}_K Z\|_2^2$  has a  $\chi_{(K)}^2$  distribution, so that for any  $c > 1$ , there is a  $d > 0$  such that

$$\mathbb{P}(\|\tilde{P}_K Z\|_2 \leq \sqrt{K}(1 + d\sqrt{\log n})) \geq 1 - n^{-c}. \quad (51)$$

Also note that for  $1 \leq k \leq K$ ,

$$\mathbf{X}\tilde{\mathbf{u}}_k = \left(\sqrt{\lambda_k} \langle \mathbf{u}_k, \tilde{\mathbf{u}}_k \rangle \mathbf{v}_k + \sigma_1 \mathbf{E}\tilde{\mathbf{u}}_k\right) + \sum_{l \neq k}^M \sqrt{\lambda_l} \langle \mathbf{u}_l, \tilde{\mathbf{u}}_k \rangle \mathbf{v}_l,$$

so that, by *Proposition 1*, (48), (49) and the fact that  $\bar{\delta}_n = o(\frac{\bar{q}_n}{n \log n})$  (by assumption **A5**), there is a constant  $\gamma_3 = \gamma_3(\sigma_1, \lambda_1, \dots, \lambda_M)$ , such that for any  $c > 1$ , there is a  $d > 0$  such that for  $n$  large enough,

$$\mathbb{P}\left(\max_{1 \leq k \leq K} \frac{1}{\sqrt{n}} \|\mathbf{X}\tilde{\mathbf{u}}_k\|_2 \leq \ell_1(1 + d\gamma_3 \sqrt{\frac{\bar{q}_n \vee \log n}{n}})\right) \geq 1 - n^{-c}. \quad (52)$$

The proof is now completed by using (14) and applying *Proposition 1* together with (48), (49), (50), (52) and **A5**, and checking that  $\|R_n\|_2$  is of a smaller asymptotic order.

**Proof of Corollary 1 :** By *Theorem 1* and *Theorem 2*, without loss of generality, can restrict attention to the set  $\{\widehat{\mathcal{A}}_{\widehat{\beta}, \mu_n} = \mathcal{A}\}$ . On this set

$$\|\widehat{\theta}_{\mathcal{A}}^{\widehat{\mathcal{A}}_{\widehat{\beta}, \mu_n}, OLS} - \theta_{\mathcal{A}}\|_2 = \|(\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1} \mathbf{X}_{\mathcal{A}}^T \widetilde{Y} - \theta_{\mathcal{A}}\|_2 = \|(\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1} \mathbf{X}_{\mathcal{A}}^T \Delta\|_2 \leq \|(\mathbf{X}_{\mathcal{A}}^T \mathbf{X}_{\mathcal{A}})^{-1}\|^{1/2} \|\Delta\|_2.$$

Now use *Lemma A.1* and *Proposition 4* to complete the proof.

**Proof of Proposition 5 :** We first show that if  $\mu_n = o(n^{-(1-\alpha)/2})$  then

$$\mathbb{P}(\exists j \in \mathcal{A}^c : |\widehat{\theta}_j^{LASSO, \mu_n}| \neq 0) \geq 1 - n^{-c_8}. \quad (53)$$

We show this by showing that, if  $\widehat{\theta}^{\mathcal{A}, L, \mu}$  and  $\widehat{\theta}^{\mathcal{A}_+, L, \mu}$  denote the solutions of the restricted LASSO problems

$$\widehat{\theta}^{\mathcal{A}, L, \mu} = \arg \min_{\zeta \in \mathbb{R}^p : \zeta_{\mathcal{A}^c} = 0} \frac{1}{n} \|Y - \mathbf{X}\zeta\|_2^2 + \mu \|\zeta\|_1, \quad (54)$$

and

$$\widehat{\theta}^{\mathcal{A}_+, L, \mu} = \arg \min_{\zeta \in \mathbb{R}^p : \zeta_{\mathcal{A}_+^c} = 0} \frac{1}{n} \|Y - \mathbf{X}\zeta\|_2^2 + \mu \|\zeta\|_1, \quad (55)$$

respectively, then  $\mathbb{P}(\max_{j \in \mathcal{A}^c} |\overline{G}_j(\widehat{\theta}^{\mathcal{A}, L, \mu_n})| > \mu_n) \geq 1 - n^{-c_8}$  provided  $\mu_n = o(n^{-(1-\alpha)/2})$ . Here  $\overline{G}_j(\zeta) = -\frac{2}{n} \langle Y - \mathbf{X}\zeta, X_j \rangle$ . On the set  $\{\widehat{\mathcal{A}}_{\mu_n}^{LASSO} = \mathcal{A}\}$  we must have  $\widehat{\theta}^{LASSO, \mu_n} = \widehat{\theta}^{\mathcal{A}, L, \mu_n}$  and  $\widehat{\theta}_j^{\mathcal{A}, L, \mu_n} \neq 0$  for all  $j \in \mathcal{A}$ , and  $\widehat{\theta}_j^{\mathcal{A}_+, L, \mu_n} = 0$

for all  $j \in \mathcal{A}^c$ . Therefore, by the analogue of *Lemma 1* with  $Y$  replacing  $\tilde{Y}$ , such a vector cannot be a LASSO solution of the unrestricted problem (37).

In the second stage we show that if  $\mu_n \geq C_3 n^{-(1-\alpha)/2}$  for some  $C_3 > 0$ , then the vector  $\hat{\theta}^{\mathcal{A}_+, L, \mu_n}$  is a solution of (54) with probability at least  $1 - O(n^{-c_8})$ , for some  $c_8 > 1$ . However, by definition  $\hat{\theta}_j^{\mathcal{A}_+, L, \mu_n} = 0$  for all  $j \in \mathcal{A}_-$ . Hence this proves (36) when  $\mu_n^{-1} = O(n^{(1-\alpha)/2})$ . Combining the two parts the proposition follows.

For the first part, observe that  $\{X_l : l \in \mathcal{D}^c\}$  are i.i.d.  $N(0, \sigma_0^2)$  and are independent of  $Y$  and  $\{X_l : l \in \mathcal{A}\}$ . Define  $R^{\mathcal{A}, L, \mu} = Y - \mathbf{X}_{\mathcal{A}} \hat{\theta}_{\mathcal{A}}^{\mathcal{A}, L, \mu}$ . Then  $\|R^{\mathcal{A}, L, \mu}\|_2^2 \geq \|Y - \mathbf{X}_{\mathcal{A}} \hat{\theta}_{\mathcal{A}}^{\mathcal{A}, OLS}\|_2^2$ , by definition of the restricted LASSO estimator. Moreover,  $\bar{G}_j(\hat{\theta}^{\mathcal{A}, L, \mu}) = -\frac{2}{n} X_j^T R^{\mathcal{A}, L, \mu}$ , so that

$$\mathbb{P}(\max_{j \in \mathcal{A}^c} |\bar{G}_j(\hat{\theta}^{\mathcal{A}, L, \mu_n})| > \mu_n) \leq \mathbb{P}(\max_{j \in \mathcal{D}^c} |\frac{2}{n} X_j^T R^{\mathcal{A}, L, \mu_n}| > \mu_n). \quad (56)$$

Now we can use the same argument as in the proof of Lemma 6 in Meinshausen (2005) to reach the conclusion, since  $p_n - |\mathcal{D}_n| = p_n - |\mathcal{B}_n| = p_n - \bar{q}_n \geq (1 - \epsilon)p_n$  for any fixed  $\epsilon > 0$  for sufficiently large  $n$ .

Next, suppose that  $\mu_n \geq C_3 n^{-(1-\alpha)/2}$  for some  $C_3 > 0$ . Without loss of generality, suppose that  $\mathcal{A}_+$  is nonempty. Then by *Lemma 1* (with  $Y$  replacing  $\tilde{Y}$ ),  $\frac{2}{n} \mathbf{X}_{\mathcal{A}_+}^T R^{\mathcal{A}_+, L, \mu_n} = \mu_n \hat{s}_{\mathcal{A}_+}$ , where  $R^{\mathcal{A}_+, L, \mu} = Y - \mathbf{X}_{\mathcal{A}_+} \hat{\theta}_{\mathcal{A}_+}^{\mathcal{A}_+, L, \mu}$ , and  $\hat{s}_j = \text{sign}(\hat{\theta}_j^{\mathcal{A}_+, L, \mu_n})$  if  $\hat{\theta}_j^{\mathcal{A}_+, L, \mu_n} \neq 0$  and  $|\hat{s}_j| \leq 1$  otherwise. Define

$X_{j \cdot \mathcal{A}_+} = X_j - \mathbf{X}_{\mathcal{A}_+} \Sigma_{\mathcal{A}_+ \mathcal{A}_+}^{-1} \Sigma_{\mathcal{A}_+ j}$  for  $j \in \mathcal{A}_+^c$ . Then we can write,

$$\begin{aligned} \frac{1}{n} X_j^T R^{\mathcal{A}_+, L, \mu_n} &= \frac{1}{n} X_{j \cdot \mathcal{A}_+}^T R^{\mathcal{A}_+, L, \mu_n} + \Sigma_{j \mathcal{A}_+} \Sigma_{\mathcal{A}_+ \mathcal{A}_+}^{-1} \left( \frac{1}{n} \mathbf{X}_{\mathcal{A}_+}^T R^{\mathcal{A}_+, L, \mu_n} \right) \\ &= \frac{1}{n} X_{j \cdot \mathcal{A}_+}^T [(I - P_{\mathbf{X}_{\mathcal{A}_+}})Y - \frac{\mu_n}{2} \mathbf{X}_{\mathcal{A}_+} \left( \frac{1}{n} \mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+} \right)^{-1} \widehat{s}_{\mathcal{A}_+}] + \frac{\mu_n}{2} \Sigma_{j \mathcal{A}_+} \Sigma_{\mathcal{A}_+ \mathcal{A}_+}^{-1} \end{aligned} \quad (57)$$

Observe that

$$X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}})Y = X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}}) \mathbf{X}_{\mathcal{A}_- \cdot \mathcal{A}_+} \theta_{\mathcal{A}_-} + \sigma_\varepsilon X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}}) \varepsilon. \quad (58)$$

Using the fact that  $X_{l \cdot \mathcal{A}_+}$  is independent of  $\mathbf{X}_{\mathcal{A}_+}$  and has i.i.d.  $N(0, \Sigma_{ll \cdot \mathcal{A}_+})$  entries, for  $l \in \mathcal{A}_+^c$ , and that  $P_{\mathbf{X}_{\mathcal{A}_+}}$  is the projection matrix on the column space of  $\mathbf{X}_{\mathcal{A}_+}$  and hence has rank  $|\mathcal{A}_+|$ , we can express the first term on the RHS of (58) as

$$X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}}) \mathbf{X}_{\mathcal{A}_- \cdot \mathcal{A}_+} \theta_{\mathcal{A}_-} = (n - |\mathcal{A}_+|) \theta_{\mathcal{A}_-}^T \Sigma_{\mathcal{A}_- j \cdot \mathcal{A}_+} + \sum_{i=1}^{n-|\mathcal{A}_+|} \widetilde{\xi}_{i,j}, \quad (59)$$

where for each  $j \in \mathcal{A}_-$ ,  $\widetilde{\xi}_{i,j}$  are i.i.d.  $N(0, \tau_j^2)$  where  $\tau_j^2 = \text{Var}(X_{1,j \cdot \mathcal{A}_+} X_{1, \mathcal{A}_- \cdot \mathcal{A}_+}^T \theta_{\mathcal{A}_-})$ .

Note that  $|\theta_{\mathcal{A}_-}^T \Sigma_{\mathcal{A}_- j \cdot \mathcal{A}_+}| \leq \Sigma_{jj}^{1/2} \|\Sigma_{\mathcal{A}_- \mathcal{A}_-}\|^{1/2} \|\theta_{\mathcal{A}_-}\|_2$  and as a consequence,

$$\tau_j^2 = \Sigma_{jj \cdot \mathcal{A}_+} \theta_{\mathcal{A}_-}^T \Sigma_{\mathcal{A}_- \mathcal{A}_- \cdot \mathcal{A}_+} \theta_{\mathcal{A}_-} + \|\theta_{\mathcal{A}_-}^T \Sigma_{\mathcal{A}_- j \cdot \mathcal{A}_+}\|_2^2 \leq 2 \Sigma_{jj} \|\Sigma_{\mathcal{A}_- \mathcal{A}_-}\|^2 \|\theta_{\mathcal{A}_-}\|_2^2.$$

Since by assumption  $\|\theta_{\mathcal{A}_-}\|_2 = o(\mu_n)$ , using Bernstein's inequality we get,

for any  $\epsilon > 0$ , for sufficiently large  $n$

$$\mathbb{P}\left(\max_{j \in \mathcal{A}_-} \left| \frac{2}{n} X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}}) \mathbf{X}_{\mathcal{A}_- \cdot \mathcal{A}_+} \theta_{\mathcal{A}_-} \right| \leq \epsilon \mu_n\right) \geq 1 - n^{-c_9}. \quad (60)$$

Since  $X_{j \cdot \mathcal{A}_+}$  and  $\varepsilon$  are independent and independent of  $\mathbf{X}_{\mathcal{A}_+}$  another application of Bernstein's inequality, after similar considerations as above, implies that

$$\mathbb{P}(\max_{j \in \mathcal{A}_-} |\frac{2}{n} \sigma_\varepsilon X_{j \cdot \mathcal{A}_+}^T (I - P_{\mathbf{X}_{\mathcal{A}_+}}) \varepsilon| \leq \epsilon \mu_n) \geq 1 - n^{-c_9}. \quad (61)$$

Next, observe that for each fixed  $s_{\mathcal{A}_+}$  with entries either  $-1$  or  $+1$ , conditionally on  $\mathbf{X}_{\mathcal{A}_+}$ ,

$$X_{j \cdot \mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+} (\mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1} \sim N(0, \Sigma_{jj \cdot \mathcal{A}_+} (\mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1}).$$

Therefore, noticing that  $\|\widehat{s}_{\mathcal{A}_+}\|_2 \leq \sqrt{q_n}$ , so that

$$|X_{j \cdot \mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+} (\mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1} \widehat{s}_{\mathcal{A}_+}| \leq \sqrt{q_n} \|X_{j \cdot \mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+} (\mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1}\|_2,$$

the term appearing on the LHS of above is (conditionally on  $\mathbf{X}_{\mathcal{A}_+}$ ) stochastically bounded by  $\sqrt{\frac{q_n}{n}} \Sigma_{jj}^{1/2} \|(\frac{1}{n} \mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1}\|^{1/2} \chi_{(|\mathcal{A}_+|)}^2$  where  $\chi_{(k)}^2$  denotes a Chi-square random variable with  $k$  degrees of freedom. Now, using the second inequality of (44), with  $\mathcal{A}_+$  replacing  $\mathcal{A}$ , recalling *Lemma A.1* and the fact that for some  $d_9 > 0$ , for all  $k \leq q_n = o(n)$ ,  $\mathbb{P}(\chi_{(k)}^2 > k(1 + d_9 \log n)) \leq \frac{1}{2} n^{-c_9 - 1}$ , we conclude that

$$\mathbb{P}(\max_{j \in \mathcal{A}_-} |\frac{2}{n} X_{j \cdot \mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+} (\frac{1}{n} \mathbf{X}_{\mathcal{A}_+}^T \mathbf{X}_{\mathcal{A}_+})^{-1} \widehat{s}_{\mathcal{A}_+}| \leq t_n) \geq 1 - n^{-c_9}, \quad (62)$$

for some  $t_n \leq C_4 \sigma_0^{-1} \Sigma_{jj}^{1/2} \frac{q_n \sqrt{\log n}}{\sqrt{n}}$ , for some constant  $C_4 > 0$  (since  $\|\Sigma_{\mathcal{A}_+ \mathcal{A}_+}^{-1}\| \leq \sigma_0^{-2}$ ). By assumption **A3**,  $t_n < \epsilon$  for  $n$  large enough.

Combining (60), (61) and (62) and taking  $\epsilon > 0$  small enough so that  $3\epsilon < (1 - \delta_+)$ , and using the fact that by assumption,

$$\max_{j \in \mathcal{A}_-} |\Sigma_{j\mathcal{A}_+} \Sigma_{\mathcal{A}_+\mathcal{A}_+}^{-1} \widehat{s}_{\mathcal{A}_+}| \leq \max_{j \in \mathcal{A}_-} \|\Sigma_{j\mathcal{A}_+} \Sigma_{\mathcal{A}_+\mathcal{A}_+}^{-1}\|_1 \leq \delta_+ < 1$$

and finally putting all the pieces together to bound the RHS of (57), we conclude that there is a  $c_8 > 1$  such that for large enough  $n$ ,

$$\mathbb{P}(\max_{j \in \mathcal{A}_-} |\overline{G}_j(\widehat{\theta}^{\mathcal{A}_+, L, \mu_n})| < \mu_n) \geq 1 - n^{-c_8}.$$

This concludes the proof by the second part of *Lemma 1*, now applied to the pair of sets  $(\mathcal{A}_+, \mathcal{A})$  rather than  $(\mathcal{A}, \mathcal{P})$  and with  $Y$  as response.

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